COMPTES RENDUS

(DOKLADY)

DE L'ACADÉMIE DES SCIENCES DE L'URSS

NOUVELLE SÉRIE

1937

IMPERIAL MYCOLOGICAL INSTITUTE
21 JULY 1937

VOLUME XV

№ 6—7

ÉDITION DE L'ACADÉMIE DES SCIENCES DE L'URSS
MOSCOU

AVIS AUX AUTEURS

Les «Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS» paraissent tous les dix jours, neuf numéros formant un volume d'environ 500 pages pourvu d'une table des matières.

Les «Comptes Rendus de l'Académie des Sciences de l'URSS» publient des communications courtes précisant l'essentiel des travaux plus volumineux en préparation. Ces articles, ne dépassant pas comme règle dix mille signes typographiques, sont donnés en l'une des trois langues: français, anglais, allemand. Ils embrassent les sciences mathématiques, physiques, naturelles et techniques, reproduisant les résultats de recherches scientifiques actuellement en cours ou à peine achevées qui représentent les plus récentes données de l'investigation scientifique.

Les auteurs ont droit chacun à 25 extraits gratuits.

Les manuscrits doivent être adressés à la Rédaction des «Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS»—Moscou, oulitza Koujbysheva, 8.

5 Juin

COMPTES RENDUS

(DOKLADY)

DE L'ACADÉMIE DES SCIENCES DE L'URSS

COMITÉ DE RÉDACTION

A. Borissiak, de l'Académie, **A. Fersman,** de l'Académie, **A. Frumkin,** de l'Académie, **A. Richter,** de l'Académie, **S. Wawilow,** de l'Académie

NOUVELLE SÉRIE

5-me année

Paraissant tous les dix jours

1937

VOLUME XV, № 6—7

TABLE DES MATIÈRES

Pages

MATHÉMATIQUES

I. M. Vinogradow, Member of the Academy. Representation of an odd number as a sum of three primes 291
D. Menšov (Menchoff). Sur les séries de fonctions orthogonales bornées dans leur ensemble 295

ASTRONOMIE

A. A. Ivanov, membre correspondant de l'Académie. La plus probable orbite de la petite planète (122) Gerda d'après les observations en 34 oppositions de 1872 à 1934 301

MÉCANIQUE

M. Kurenskij (Kourensky). Formules fondamentales pour le calcul des éléments de la trajectoire du centre de gravité d'un projectile . . . 307

PHYSIQUE TECHNIQUE

V. Furdujev. A Method of Acoustical Design of Rooms Equipped with Loud Speakers 313
S. Lifshitz. Experimental Investigations of Reverberation Optimum for Different Frequencies 315

PHYSIQUE DU GLOBE

W. W. Shoulejkin, Corresponding Member of the Academy. The Principles of the Monsoon Theory 319

PHYSIQUE

L. A. Tumermann and **V. Šymanovskij.** A Fluorometer Based on the Effect of Debye and Sears 323
K. S. Ljalikov. Vérification expérimentale de la formule de Thomson . 327

divided also into two sets (d') and (d''). The first consists of the numbers d satisfying the condition that all prime divisors are

$$\leqslant n^{3h},$$

the second consists of all the remaining numbers d.

The sets (d_0) and (d_1) are correspondingly divided into sets (d_0'), (d_0'') and (d_1'), (d_1'').

Lemma 1. Let (x) and (y) denote two sets of increasing positive integral numbers;

$$1 < U_0 < U_1 \leqslant N_1 \leqslant N;$$

m is integral and > 0.

$$\alpha = \frac{a}{q} + \frac{\theta}{q\tau};\ (a, q) = 1;\ 0 < q \leqslant \tau;\ m = m_1\delta;\ q = q_1\delta;\ \delta = (m, q);$$

$$T = \sum_x \sum_y e^{2\pi i\alpha mxy}$$

where x runs over the numbers (x) satisfying the condition

$$U_0 < x \leqslant U_1$$

and y for a given x runs over the numbers (y) satisfying the condition

$$0 < y \leqslant \frac{N_1}{x}.$$

Then we have

$$T \ll N_1 n \sqrt{\frac{n}{U_0} + \frac{U_1}{N_1} + \frac{q_1 n}{N_1} + \frac{1}{q_1} + \frac{m_1}{\tau}}.$$

Theorem 1. Let

$$\alpha = \frac{a}{q} + \frac{\theta}{q\tau};\ (a, q) = 1;\ n^{3h} \leqslant q \leqslant \tau.$$

Then we have

$$S = \sum_{p \leqslant N} e^{2\pi i\alpha p} \ll Nn^{2-h}.$$

Proof. We have

$$S = \sum_{(d)} \mu(d)\, S_d + O(\sqrt{N});\ S_d = \sum_{m=1}^{\frac{N}{d}} e^{2\pi i\alpha md}. \tag{1}$$

Hence we find

$$S = \sum_{d > \tau_1} \mu(d)\, S_d + O(Nn^{-h+1}) =$$

$$= T_0 - T_1 + O(Nn^{-h+1});\quad T_0 = \sum_{(d_0)} S_d;\quad T_1 = \sum_{(d_1)} S_d, \tag{2}$$

where d runs over the values $> \tau_1$. We estimate only T_0 as the estimation of T_1 is performed in the same way.

By change of the order of the summation we get

$$T_0 = \sum_m T(m);\quad T(m) = \sum_d e^{2\pi i\alpha md}, \tag{3}$$

where m runs over the numbers

$$m = 1, \ \ldots, \ [n^h]$$

and d for every given m runs over the numbers

$$\tau_1 < d \leqslant \frac{N}{m}\,.$$

Further we find

$$T(m) = T''(m) + O\Big(\frac{N}{m}\,n^{-h}\Big) \tag{4}$$

where $T''(m)$ consists only of those terms of the sum $T(m)$ which correspond to the values of d from the set (d_0''). For, the part $T'(m)$ of $T(m)$ corresponding to the values of d from d_0' does not exceed the number of those terms of the set (d') which are not exceeding $\frac{N}{m}$. But the order of this number is significantly less than

$$\frac{N}{m}\,n^{-h}.$$

Now if d belongs to (d_0'') and k is the number of the prime divisors of d exceeding n^{3h}, then

$$k < n.$$

Therefore

$$T''(m) = \sum_{k<n} T_k(m), \tag{5}$$

where $T_k(m)$ consists of the terms of $T''(m)$ with d containing exactly k prime divisors $> n^{3h}$. Further we find

$$T_k(m) = \frac{1}{k}\,T_{k0}(m) + O\Big(\frac{N}{mk}\,n^{-3h}\Big), \tag{6}$$

where

$$T_{k0}(m) = \sum_u \sum_v e^{2\pi i \alpha m u v},$$

and u runs over the primes $\geqslant n^{3h}$ belonging to (d) and v for a given u runs over the numbers belonging to (d_1) and satisfying the condition

$$\frac{\tau_1}{u} < v \leqslant \frac{N}{mu}\,.$$

To the sum $T_{k0}(m)$ we can directly apply the lemma 1. We obtain

$$T_{k0}(m) \ll N\,\frac{n^{\frac{3}{2}-\frac{3}{2}h}}{\sqrt{m}}\,,$$

and then from (6), (5), (4), (3) and (2) we deduce the theorem 1.

Theorem 2. For the number I_N of the representations of N in the form

$$N = p_1 + p_2 + p_3$$

we have the formula

$$I_N = RS + O(N^2 n^{-c}),$$

where c is an arbitrary large constant > 3 and

$$S = \sum_{q=1}^{\infty} \frac{\mu(q)}{q_1^3} \sum_{\substack{0 \leqslant a < q \\ (a,\,q)=1}} e^{2\pi i \frac{a}{q} N}\,; \quad R = \frac{N^2}{2n^3}(1+\lambda); \quad \lim_{N\to\infty} \lambda = 0; \quad q_1 = \varphi(q).$$

Proof. a) We have

$$I_N = \int_0^1 S_\alpha^3 e^{-2\pi i \alpha N}\, d\alpha; \quad S_\alpha = \sum_{p \leqslant N} e^{2\pi i \alpha p}.$$

The interval of integration we divide into intervals of two classes

1. $\alpha = \frac{a}{q} + z; \quad (a, q) = 1; \quad 0 < q \leqslant n^{3h}; \quad -\frac{1}{\tau} \leqslant z \leqslant \frac{1}{\tau}.$

2. The remaining intervals; for them

$$\alpha = \frac{a}{q} + z; \quad (a, q) = 1; \quad n^{3h} < q \leqslant \tau; \quad |z| \leqslant \frac{1}{q\tau}.$$

Correspondingly this division into intervals we have

$$I_N = I_{N1} + I_{N2}. \tag{7}$$

b) Applying the theorem 1 we find

$$I_{N2} \ll Nn^{2-h} \int_0^1 |S_\alpha|^2\, d\alpha \ll Nn^{2-h} \int_0^1 \sum_{p \leqslant N} \sum_{p_1 \leqslant N} e^{2\pi i \alpha (p - p_1)}\, d\alpha \ll$$

$$\ll Nn^{2-h} \frac{N}{n} \ll N^2 n^{1-h}. \tag{8}$$

c) The calculation of I_{N1} we perform without difficulty. This is done in the same way as in the Waring's problem. But here we use the new theorem on the distribution of primes in an arithmetical progression. If α belongs to the interval of the first class then we find

$$S_\alpha = \frac{\mu(q)}{q_1} V(z) + O(Nn^{-h_1}); \quad V(z) = \int_2^N \frac{e^{2\pi i z x}}{\log x}\, dx.$$

Hence the part of I_{N1} corresponding the given fraction $\frac{a}{q}$ is represented in the form

$$R \frac{\mu(q)}{q_1^3} e^{-2\pi i \frac{a}{q} N} + O(Nn^{-h_2}); \quad R = \int_{-\frac{1}{\tau}}^{\frac{1}{\tau}} [V(z)]^3 e^{-2\pi i z N}\, dz,$$

where R can be expressed in the form

$$R = \frac{N^2}{2n^3}(1 + \lambda); \quad \lim_{N \to \infty} \lambda = 0.$$

Hence we find without difficulty:

$$I_{N1} = RS + O(N^2 n^{-h_3})$$

and by (7) and (8) the theorem follows directly.

V. Stekloff Institute of Mathematics.
Academy of Sciences of the USSR.
Moscow.

Received
19.V.1937.

REFERENCES

[1] И. М. Виноградов, ДАН, III, 1; IV, 4 (1934). [2] A. Walfitz, Math. ZS., 40 (1936).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

MATHÉMATIQUES

SUR LES SÉRIES DE FONCTIONS ORTHOGONALES BORNÉES DANS LEUR ENSEMBLE

Par D. MENŠOV (MENCHOFF)

(Présenté par I. M. Vinogradow, de l'Académie, le 10. IV. 1937)

Soit

$$\varphi_1(x),\ \varphi_2(x),\ \ldots,\ \varphi_n(x)\ \ldots \tag{1}$$

un système normé de fonctions orthogonales dans l'intervalle (0,1), donc

$$\int_0^1 \varphi_n(x)\,\varphi_m(x)\,dx = \begin{cases} 0, & n \neq m, \\ 1, & n = m. \end{cases} \tag{2}$$

On sait que la série

$$\sum_{n=1}^{\infty} c_n \varphi_n(x) \tag{3}$$

converge presque partout dans (0,1), si la série

$$\sum_{n=1}^{\infty} (\lg n)^2 c_n^2 \tag{4}$$

converge, les c_n étant des constantes réelles (1). D'ailleurs, dans la série (4) on ne peut pas remplacer la fonction $(\lg n)^2$ par une autre fonction positive $W(n)$ croissant moins vite, c'est-à-dire telle que

$$\lim_{n\to\infty} \frac{W(n)}{(\lg n)^2} = 0. \tag{5}$$

En effet, si $W(n)$ est une fonction positive quelconque vérifiant la condition (5), on peut toujours déterminer un système normé de fonctions orthogonales (1) et une suite de constantes réelles c_n, $n = 1, 2, 3, \ldots$, telles que la série (3) diverge partout dans (0,1), tandis que la série

$$\sum_{n=1}^{\infty} W(n)\, c_n^2 \tag{6}$$

converge (2).

Nous dirons que les fonctions $\varphi_n(x)$, $n = 1, 2, 3, \ldots$, sont b o r n é e s d a n s l e u r e n s e m b l e dans l'intervalle (0,1), s'il existe un nombre positif M ne dépendant ni de n ni de x tel que l'inégalité

$$|\varphi_n(x)| < M$$

est vérifiée pour toutes les valeurs de x dans (0,1) et pour toutes les valeurs entières et positives de n. On sait qu'à toute fonction positive $W(n)$ vérifiant la condition $\lim_{n\to\infty} \frac{W(n)}{\lg n} = 0$ il correspond un système normé (1) de fonctions orthogonales bornées dans leur ensemble et une suite de constantes réelles c_n, $n = 1, 2, 3, \ldots$, telles que la série (3) diverge partout dans (0,1), tandis que la série (6) converge (3). On peut se demander si la convergence de la série $\sum_{n=1}^{\infty} \lg n \cdot c_n^2$ entraîne la convergence presque partout dans (0,1) de la série (3) dans le cas où les fonctions $\varphi_n(x)$, $n=1, 2, 3, \ldots$, sont bornées dans leur ensemble sur l'intervalle (0,1). La réponse à cette question est négative. En effet, on peut démontrer le t h é o r è m e.

Pour toute fonction positive $W(n)$ vérifiant la condition (5) *on peut déterminer un système normé* (1) *de fonctions orthogonales bornées dans leur ensemble dans* (0,1) *et une suite de constantes réelles c_n, $n = 1, 2, 3, \ldots$, telles que la série* (3) *diverge partout dans* (0,1), *tandis que la série* (6) *converge.*

Il suit des raisonnements que l'on fait en considérant des problèmes analogues (4) que le théorème en question est une conséquence du

l e m m e. *Pour chaque entier positif m suffisamment grand, on peut déterminer les fonctions $\psi_k(x)$ et les constantes réelles A_k, $1 \leqslant k \leqslant m$, qui vérifient les conditions suivantes:*

1. *Les $\psi_k(x)$, $1 \leqslant k \leqslant m$, constituent un système normé de fonctions orthogonales dans un intervalle (a, b) qui ne dépend pas de m.*

2. *Il existe un intervalle (a', b'), $a' < b'$, situé dans (a, b) et ne dépendant pas de m, tel que*

$$\sum_{k=1}^{l(x)} A_k \psi_k(x) > C$$

pour toutes les valeurs de x dans (a', b'), où $l(x)$ dépend en général de x, $1 \leqslant l(x) \leqslant m$, et C est une constante absolue positive.

3. $$\sum_{k=1}^{m} A_k^2 < \frac{C'}{(\lg m)^2},$$

où C' est une constante absolue.

4. $$|\psi_k(x)| < C'' \qquad (a \leqslant x \leqslant b,\ 1 \leqslant k \leqslant m),$$

où C'' est aussi une constante absolue.

5. *Les fonctions $\psi_k(x)$, $1 \leqslant k \leqslant m$, sont constantes dans chacun des intervalles en nombre fini obtenus par une subdivision convenable de l'intervalle (a, b).*

Il est clair que le lemme sera démontré si l'on définit les fonctions $\psi_k(x)$ et les constantes A_k, vérifiant les conditions 1, 2, 3, 4 et 5, pour $m = \nu^2$, où ν est un nombre entier supérieur à 3. Nous commençons par définir quelques fonctions auxiliaires.

Posons

$$\left.\begin{aligned} \omega_k(x) &= \frac{\nu}{\nu^2 x - k - \nu} \qquad \left(-\infty < x < \frac{k}{\nu^2}\right), \\ \omega_k(x) &= \frac{\nu}{\nu^2 x - k + \nu} \qquad \left(\frac{k}{\nu^2} \leqslant x < +\infty\right). \end{aligned}\right\} \tag{8}$$

En supposant que $\nu > 3$, on peut démontrer que les fonctions $\omega_k(x)$, $1 \leqslant k \leqslant \nu^2$, possèdent les propriétés suivantes:

α) $$\left|\int_{-1}^{2} \omega_k(x)\,\omega_l(x)\,dx\right| < \frac{4\nu}{p(p+2\nu)} \lg\left(1+\frac{p}{\nu}\right) + \frac{2}{\nu^2}$$

$$(k \neq l,\ 1 \leqslant k \leqslant \nu^2,\ 1 \leqslant l \leqslant \nu^2,\ p = |k-l|);$$

β) $$|\omega_k(x)| \leqslant 1 \qquad (-\infty < x < +\infty,\ 1 \leqslant k \leqslant \nu^2);$$

γ) $$\sum_{k=1}^{l(x)} \omega_k(x) > \frac{1}{2}\nu \lg \nu \qquad \left(\frac{1}{2} \leqslant x \leqslant 1\right),$$

où $l(x)$ dépend en général de x et $1 \leqslant l(x) \leqslant \nu^2$.

Nous remplacerons les fonctions $\omega_k(x)$ par des fonctions $\chi_k(x)$, $1 \leqslant k \leqslant \nu^2$, définies dans l'intervalle $(-1, 2)$ et vérifiant les conditions:

a) $$|\chi_k(x) - \omega_k(x)| < \frac{1}{\nu^2} \qquad (-1 \leqslant x \leqslant 2,\ 1 \leqslant k \leqslant \nu^2);$$

b) les fonctions $\chi_k(x)$, $1 \leqslant k \leqslant \nu^2$, sont constantes dans chacun des intervalles en nombre fini obtenus par une subdivision convenable de l'intervalle $(-1,2)$.

On démontre sans peine que les fonctions $\chi_k(x)$ possèdent les propriétés suivantes:

c) $$|\alpha_{kl}| < \gamma_p \qquad (k \neq l,\ 1 \leqslant k \leqslant \nu^2,\ 1 \leqslant l \leqslant \nu^2),$$

où

$$\alpha_{kl} = \int_{-1}^{2} \chi_k(x)\,\chi_l(x)\,dx, \quad \gamma_p = \frac{4\nu}{p(p+2\nu)} \lg\left(1+\frac{p}{\nu}\right) + \frac{9}{\nu^2}, \quad p = |k-l|; \tag{9}$$

d) $$|\chi_k(x)| < 2 \qquad (-1 \leqslant x \leqslant 2,\ 1 \leqslant k \leqslant \nu^2);$$

e) $$\sum_{k=1}^{l(x)} \chi_k(x) > \frac{1}{4}\nu \lg \nu \qquad \left(\frac{1}{2} \leqslant x \leqslant 1\right),$$

où $l(x)$ dépend en général de x et $1 \leqslant l(x) \leqslant \nu^2$.

En posant

$$c_0 = 2, \quad c_p = 2\left(1 + \sum_{s=1}^{p} \gamma_s\right) \qquad (1 \leqslant p < \nu^2), \tag{10}$$

on a évidemment

$$c_p - c_{p-1} = 2\gamma_p \qquad (1 \leqslant p < \nu^2). \tag{11}$$

D'ailleurs, on peut démontrer que

$$c_{\nu^2-1} < b, \tag{12}$$

où

$$b = 4\left[5 + 2\int_0^{\infty} \frac{\lg(1+z)}{z(z+2)}\,dz\right], \tag{13}$$

c'est-à-dire b possède une valeur finie qui ne dépend pas de ν.

Nous avons déjà déterminé dans l'intervalle $(-1,2)$ un système de ν^2 fonctions $\chi_k(x)$ qui possèdent les propriétés a), b), c), d) et e).

En tenant compte de la propriété c) de ces fonctions et de l'égalité (11), nous pouvons déterminer dans l'intervalle $(2, b)$ un système de ν^2 fonctions que nous désignerons aussi par $\gamma_k(x)$, $1 \leqslant k \leqslant \nu^2$, et qui possèdent les propriétés suivantes:

f) Les fonctions $\gamma_k(x)$ ne prennent que deux valeurs $+1$ et -1 pour $2 < x \leqslant b$ et sont constantes dans chacun des intervalles en nombre fini obtenus par une subdivision convenable de l'intervalle $(2, b)$.

g) $$\int_{c_{p-1}}^{c_p} \gamma_k(x)\,\gamma_l(x)\,dx = \begin{cases} -\alpha_{kl}, & \text{si } |k-l| = p, \\ 0, & \text{si } |k-l| \neq p, \end{cases}$$

$$(k \neq l,\ 1 \leqslant k \leqslant \nu^2,\ 1 \leqslant l \leqslant \nu^2).$$

h) $$\int_{c_{\nu^2-1}}^{b} \gamma_k(x)\,\gamma_l(x)\,dx = 0 \qquad (k \neq l)^*.$$

En posant $a = -1$ et en tenant compte de l'égalité (9) et des propriétés g) et h) des fonctions $\gamma_k(x)$, on obtient immédiatement

$$\int_a^b \chi_k(x)\,\chi_l(x)\,dx = 0$$

$$(k \neq l,\ 1 \leqslant k \leqslant \nu^2,\ 1 \leqslant l \leqslant \nu^2). \tag{14}$$

Posons

$$\psi_k(x) = \frac{\chi_k(x)}{\sqrt{\int_a^b \chi_k^2(x)\,dx}}, \tag{15}$$

d'où il résulte

$$\int_a^b \psi_k^2(x)\,dx = 1 \qquad (1 \leqslant k \leqslant \nu^2). \tag{16}$$

En tenant compte des propriétés d) et f) des fonctions $\gamma_k(x)$, on obtient

$$2(b-a) > \int_a^b \gamma_k^2(x)\,dx > b-2. \tag{17}$$

---

* On définit les fonctions $\chi_k(x)$ possédant les propriétés f), g) et h) en se servant des raisonnements déjà employés par A. Kolmogoroff et D. Menchoff, loc. cit.

Puisque $b-2$ est une constante absolue positive, il résulte de (15), (17) et des propriétés d) et f) des fonctions $\chi_k(x)$ que

$$|\psi_k(x)| < C'' \qquad (a \leqslant x \leqslant b,\ 1 \leqslant k \leqslant \nu^2),$$

donc les fonctions $\psi_k(x)$ vérifient la condition 4 du lemme (pour $m = \nu^2$). De plus, d'après (14) et (16) les $\psi_k(x)$, $1 \leqslant k \leqslant \nu^2$ constituent un système normé de fonctions orthogonales dans (a, b), c'est-à-dire vérifient la condition 1 du lemme.

Posons

$$a' = \frac{1}{2},\ b' = 1,\ A_k = \frac{1}{\nu \lg \nu} \cdot \sqrt{\int_a^b \chi_k^2(x)\,dx}, \tag{18}$$

d'où il résulte, en vertu de (17),

$$\frac{b-2}{\nu \lg \nu} < A_k < \frac{2(b-a)}{\nu \lg \nu} \qquad (1 \leqslant k \leqslant \nu^2). \tag{19}$$

En tenant compte des égalités (15), (18), de l'inégalité (19) et des propriétés e), b), f) des fonctions $\chi_k(x)$, on peut démontrer sans peine que ces fonctions, ainsi que les constantes A_k, $1 \leqslant k \leqslant \nu^2$, vérifient les conditions 2, 3 et 5 du lemme, ce qui achève sa démonstration. Le théorème énoncé dans cet article est donc démontré *.

Institut mathématique V. A. Stekloff.
Académie des Sciences de l'URSS.
Moscou.

Manuscrit reçu
le 10. IV. 1937.

LITTÉRATURE CITÉE

[1] H. Rademacher, Math. Annalen, 87, 112 (1922). D. Menchoff, Fundamenta Mathematicae, 4, 82 (1923). [2] D. Menchoff, loc. cit. [3] A. Kolmogoroff et D. Menchoff, Math. ZS., 26, 432 (1927). [4] D. Menchoff, loc. cit. ou A. Kolmogoroff et D. Menchoff, loc. cit.

* Une démonstration plus détaillée de ce théorème sera publiée dans le Recueil Mathématique.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

ASTRONOMIE

LA PLUS PROBABLE ORBITE DE LA PETITE PLANÈTE (122) GERDA D'APRÈS LES OBSERVATIONS EN 34 OPPOSITIONS de 1872 à 1934

Par A. A. IVANOV, Membre correspondant de l'Académie

Le but de cet ouvrage est d'obtenir une orbite de la petite planète (122) Gerda qui pourrait satisfaire le mieux possible à toutes les observations de la planète depuis le moment de sa découverte, c'est-à-dire depuis 1872, jusqu'à présent. Dans cet ouvrage je me suis servi des résultats publiés dans deux de mes anciens travaux (1) et (2) ayant trait aussi à la planète (122) Gerda.

Comme éléments initials j'ai adopté ici les éléments suivants:

T 1904 Janvier 24.0, temps moyen de Berlin

$i = 1°36'34''.26$
$\Omega = 178\ 39\ 56\ .92$
$\omega = 13\ 6\ 24\ .40$
} équinoxe moyen 1904.0

$M_0 = 287°58'29''.92$, $\mu = 615''.55400$
$\varphi = 3\ 0\ 19\ .62$, $\log a = 0.507604$

Au cours de la période de 1872 à 1934 la planète Gerda fut observée en 34 oppositions, et pour chacune de ces oppositions j'ai fait une place normale que je donne ci-dessous.

№	Date	Temps moyen de Berlin	α	δ	Équinoxe	$\Delta\alpha$	$\Delta\delta$
1	1872 Août . . .	29.0	$21^h28^m54^s.10$	$-13°32'27''8$	1872.0	$-70''05$	$-22''0$
2	1873 Oct.	26.0	2 4 43 .02	+11 17 59.7	1873.0	— 41.40	— 14.8
3	1876 Avr. . . .	8.0	11 54 53 .88	+ 0 48 56.4	1876.0	+ 33.90	— 12.5
4	1877 Juin . . .	30.0	17 20 51 .94	−20 49 8.3	1877.0	+ 36.30	+ 0.6
5	1878 Sept. . . .	19.0	22 27 51 .48	− 8 55 17.0	1878.0	+ 32.70	+ 13.0
6	1885 Déc. . . .	9.0	3 54 13 .75	+18 11 41.5	1885.0	+ 57.45	+ 11.8
7	1893 Mars . . .	16.0	10 11 46 .85	+10 20 0.4	1893.0	+ 28.05	— 8.7
8	1896 Nov. . . .	5.0	1 22 30 .69	+ 7 30 29.7	1896.0	+ 35.85	+ 16.1
9	1897 Déc. . . .	25.0	6 21 31 .37	+21 3 34.9	1897.0	— 5.55	— 1.7
10	1899 Mars . . .	18.0	11 36 21 .38	+ 2 23 12.5	1899.0	— 5.85	+ 4.8
11	1901 Août . . .	15.0	22 26 21 .38	− 8 38 36.1	1901.0	+ 10.65	+ 2.5
12	1902 Nov. . . .	11.0	2 27 45 .39	+12 58 44.4	1902.0	— 22.05	— 8.8
13	1904 Janv. . . .	25.0	7 23 49 .19	+19 56 40.0	1904.0	— 76.80	+ 8.6
14	1905 Avr. . . .	7.0	12 53 32 .40	− 4 59 49.9	1905.0	— 64.35	+ 26.4
15	1906 Juil. . . .	8.0	18 25 12 .23	−21 0 46.3	1906.0	— 27.30	— 4.1
16	1907 Sept. . . .	14.0	23 17 28 .83	− 4 9 44.8	1907.0	— 18.00	— 5.6
17	1908 Nov. . . .	24.0	3 32 13 .93	+17 7 16.2	1908.0	— 52.65	— 14.5
18	1910 Févr. . . .	5.0	8 44 25 .33	+16 18 20.4	1910.0	−104.10	+ 23.0
19	1911 Mai . . .	3.0	14 7 49 .05	−11 22 29.4	1911.0	— 99.15	+ 34.2
20	1913 Oct.	10.0	0 7 23 .98	+ 0 29 0.1	1913.0	— 31.80	— 13.8
	Temps moyen de Greenwich						
21	1916 Mars . . .	16.0	9 48 49 .28	+12 14 20.9	1916.0	−144.15	+ 61.1
22	1917 Mai . . .	25.0	15 28 17 .45	−16 47 45.2	1917.0	−119.40	+ 32.7

№	Date	Temps moyen de Greenwich	α	δ	Équinoxe	$\Delta\alpha$	$\Delta\delta$
23	1918	Août . . . 1.0	$21^h\ 0^m10^s.40$	$-15°16'\ 4''6$	1918.0	$-\ 58''95$	$-\ 11''0$
24	1921	Janv. . . . 29.0	5 29 51 .03	+21 10 1.2	1921.0	−178.95	− 8.8
25	1922	Mai . . . 23.0	11 5 41 .42	+ 6 0 4.2	1922.0	−145.05	+ 53.3
26	1923	Mai 29.0	17 3 46. 44	−20 31 15.6	1923.0	− 99.60	+ 11.4
27	1924	Nov. . . . 29.0	21 57 12 .42	−12 2 45.0	1924.0	− 23.85	− 3.4
		Temps universel					
28	1927	Janv. . . . 7.0	7 14 18 .90	+20 6 17.5	1927.0	−291.00	+ 30.1
29	1928	Avr. . . . 13.0	12 33 54 .56	− 3 0 6.3	1928.0	−236.55	+ 96.0
30	1929	Juil. . . . 6.0	18 8 35 .09	−21 6 23.0	1929.0	− 99.15	+ 5.2
31	1930	Sept. . . . 19.0	22 52 54 .00	− 6 35 24.0	1925.0	− 69.45	− 39.7
32	1931	Nov. . . . 12.0	3 17 32 .00	+16 21 36.0		−170.40	− 39.9
33	1933	Févr. . . . 25.0	8 10 52 .00	+18 22 30.0		−420.45	+131.0
34	1934	Mai 13.0	13 47 59 .66	− 9 37 17.0		−240.75	+ 85.8

La détermination de la plus probable orbite de la planète Gerda fut faite d'après la méthode des moindres carrés avec utilisation du schème de Bauschinger (3).

Aux places normales donneés ci-dessus correspondent les 68 équations conditionnelles suivantes, dont les premières 34 se rapportent à l'ascension droite, et les autres 34 à la déclinaison.

	adM_0	+ $bd\mu$	+ $cd\varphi$	+ ddS	+ edp	+ jdq	= l
1	0.11921	4.17949_n	0.37598	0.13326	9.55722_n	9.33095_n	1.83318_n
2	0.00929	4.14243_n	9.59427_n	0.13020	8.82519	9.65199_n	1.60849_n
3	0.15503	4.16253_n	0.03511_n	0.12514	9.26984_n	9.70311_n	1.53016
4	0.17731	4.16517_n	0.39445	0.15869	8.90899	8.65034_n	1.53059
5	0.09326	4.06094_n	0.25451	0.11816	9.49763_n	9.56106_n	1.50926
6	0.10965	3.93179_n	0.25858_n	0.13710	9.28247	9.28931$_n$	1.73702
7	0.15882	3.76016_n	0.25806_n	0.12643	9.45950_n	9.57534_n	1.44082
8	0.06978	3.49484_n	9.60745_n	0.11153	9.00832	9.67437_n	1.55075
9	0.14979	3.49591_n	0.44752_n	0.15674	8.68314_n	7.99283	0.71427_n
10	0.17931	3.42891_n	0.02004_n	0.13818	9.29579_n	9.70820_n	0.76678_n
11	0.09189	3.03430_n	0.22559	0.12575	9.50963_n	9.55958_n	1.02239
12	0.08988	2.74303_n	0.09247_n	0.12932	9.31263	9.56457_n	0.94448_n
13	0.16166	1.37241_n	0.44961_n	0.15194	9.26951_n	8.60905_n	1.85850_n
14	**0.19803**	2.83668	9.22356	0.15166	8.56891	9.74068_n	1.80689_n
15	0.16782	3.11443	**0.46432**	0.16300	8.76829_n	7.35524	1.40628_n
16	0.08031	3.20298	9.98859	0.12249	9.28633_n	9.68440_n	1.25412_n
17	0.10839	3.35344	0.28282_n	0.14029	9.35795	9.34485_n	1.70170_n
18	0.17478	3.51637	0.39986_n	0.15033	9.48895_n	9.25226_n	1.99963
19	0.19391	3.61670	0.08605	0.14911	9.31660	9.62483_n	1.98766_n
20	0.06577	3.61340	9.17101	0.11373	8.08182_n	9.71746_n	1.50241_n
21	0.16330	3.80685	0.21272_n	0.12310	9.36811_n	9.57559_n	2.14886_n
22	0.19081	3.87738	0.34728	0.15768	9.41952	9.32972_n	2.05805_n
23	0.11526	3.84021	0.34344	0.14263	9.48778_n	9.34398_n	1.75488_n
24	0.10310	3.89296	0.40834_n	0.10675	8.80884	6.20509_n	2.22238_n
25	0.05918	3.88047	9.60421_n	0.00804	7.85356_n	9.62014_n	2.15915_n

	adM_0	+ $bd\mu$	+ $cd\varphi$	+ dds	+ edp	+ fdq	= l
26	0.18231	4.03216	0.44894	**0.16453**	9.09877	8.65068$_n$	1.96978$_n$
27	9.91048	3.78795	9.90212	9.95514	9.00351$_n$	9.47828$_n$	1.36782$_n$
28	0.17429	4.09783	0.44811$_n$	0.15888	9.22028$_n$	8.65214$_n$	2.43661$_n$
29	0.18782	4.13382	9.68708	0.13736	9.05948	9.72611$_n$	2.37331$_n$
30	0.16090	4.12849	0.46143	0.16113	8.29283$_n$	7.19468$_n$	1.96613$_n$
31	0.07050	4.05808	9.86412	0.11846	9.10520$_n$	9.68587$_n$	1.83880$_n$
32	0.10891	4.11573	0.32039$_n$	0.13992	9.42151	9.34601$_n$	2.21350$_n$
33	0.15887	4.18359	0.34674$_n$	0.12653	9.32126$_n$	9.26763$_n$	**2.60109**$_n$
34	0.17902	**4.22205**	0.19164	0.13420	9.46664	9.58334$_n$	2.37539$_n$
35	9.60752	3.66766$_n$	9.86241	9.62180	0.06929	9.84302	1.34242$_n$
36	9.62581	3.66915$_n$	9.08825$_n$	9.65677	9.29863$_n$	0.12543	1.17026$_n$
37	9.75735$_n$	3.76497	9.64419	9.72750$_n$	9.66686$_n$	0.10013$_n$	1.09691$_n$
38	8.97740$_n$	2.96724	9.16482$_n$	8.95576$_n$	0.10314	9.84449$_n$	9.77815
39	9.65470	3.62236$_n$	9.81559	9.67964	9.93628	9.99971	1.11394
40	9.40429	3.22738$_n$	9.53960$_n$	9.43257	9.98472$_n$	9.99156	1.07188
41	9.70404$_n$	3.30534	9.80275	9.67163$_n$	9.91441$_n$	0.03025$_n$	0.93952$_n$
42	9.63877	3.06484$_n$	9.14292$_n$	9.68064	9.43803$_n$	0.10406	1.20683
43	8.69176$_n$	2.02202	8.99739	8.69117$_n$	0.14787$_n$	9.45756	0.23045$_n$
44	9.77950$_n$	3.02906	9.61982	9.73837$_n$	9.69558$_n$	0.10799$_n$	0.68124
45	9.65069	2.59036$_n$	9.77905	9.68491	9.94979	9.99974	0.39794
46	9.58291	2.24341$_n$	9.57338$_n$	9.62272	9.81860$_n$	0.07054	1.33217$_n$
47	9.29237$_n$	9.16040$_n$	9.57491	9.28093$_n$	0.14313$_n$	9.48267$_n$	0.93450
48	9.77856$_n$	2.41937$_n$	8.79615$_n$	9.73219$_n$	8.96960	**0.14137**$_n$	1.42160
49	8.77279	1.75183	9.07892	8.77503	**0.16325**	8.75020$_n$	0.61278$_n$
50	9.67437	2.79745	9.59996	9.71660	9.69222	0.09029	0.74819$_n$
51	9.46832	2.71048	9.63332$_n$	9.50107	9.99641$_n$	9.98331	1.16137$_n$
52	9.57736$_n$	2.92050$_n$	9.79697	9.55217$_n$	0.08771$_n$	9.85102$_n$	1.36173
53	9.71598$_n$	3.13772$_n$	9.59257$_n$	9.67085$_n$	9.79423	0.10246$_n$	1.53403
54	9.66815	3.21546	8.80912	9.71610	8.47913	0.11477	1.13988$_n$
55	9.68071$_n$	3.32431$_n$	9.72932	9.64048$_n$	9.85089$_n$	0.05837$_n$	1.78604
56	9.56202$_n$	3.24756$_n$	9.70715$_n$	9.52788$_n$	0.04801	9.95821$_n$	1.51455
57	9.54945	3.27514	9.77253	9.57759	0.05241	9.90861	1.05308$_n$
58	8.77166	2.55837	9.01870$_n$	8.78021	0.11759$_n$	7.51384	0.94448$_n$
59	9.64343$_n$	3.46431$_n$	9.22451	9.59231$_n$	8.26500$_n$	0.03158$_n$	1.72673
60	9.14833$_n$	2.99639$_n$	9.40107$_n$	9.12735$_n$	0.13922	9.69113$_n$	1.05690
61	9.43486	3.31194	9.42500	9.47986	9.47553	9.95030	0.53148$_n$
62	9.25304$_n$	3.17773$_n$	9.51822	9.23517$_n$	0.14354$_n$	9.57540$_n$	1.47857
63	9.78469$_n$	3.73053$_n$	9.26249$_n$	9.73415$_n$	9.46213	0.12876$_n$	1.98227
64	8.29490	2.27199	8.60287	8.31776	0.16057	9.06242	0.71600
65	9.65144	3.63905	9.44349	9.69941	9.52432	0.10499	1.59879$_n$
66	9.50606	3.51248	9.71139$_n$	9.53786	0.02393$_n$	9.94843	1.60097$_n$
67	9.47392$_n$	3.49886$_n$	9.65665	9.44110$_n$	0.00871$_n$	9.95508$_n$	2.11727
68	9.72334$_n$	3.76610$_n$	9.72412$_n$	9.67798$_n$	9.92093	0.03763$_n$	1.93349

Pour calculer les coefficients des équations conditionnelles je me suis servi de quelques formules de contrôle.

A savoir, si l'on joint aux formules, qui définissent dans le schème de Bauschinger les valeurs de b, B, c, C les formules

$$a' \sin A = -\cos\delta \cos(\alpha - \Omega)$$

$$a' \cos A = -\sin i \sin\delta - \cos i \cos\delta \sin(\alpha - \Omega),$$

dans lesquelles $a' > 0$, on a la formule de contrôle suivante:

$$\operatorname{tg} i' = \frac{a' \sin c \sin(C - A)}{\sin b \cos B}.$$

En outre, les corrélations suivantes doivent être satisfaites:

$$F_c^\alpha \cos(B + \omega + \varphi) + F_s^\alpha \sin(B + \omega + \varphi) =$$
$$= G_c^\alpha \cos(B + \omega) + G_s^\alpha \sin(B + \omega) = \frac{a}{\rho} \sin b,$$
$$F_c^\delta \cos(C + \omega + \varphi) + F_s^\delta \sin(C + \omega + \varphi) =$$
$$= G_c^\delta \cos(C + \omega) + F_s^\delta \sin(C + \omega) = \frac{a}{\rho} \sin c,$$

$$H_c^\delta \cos\varphi + H_s^\delta \sin\varphi = \frac{r}{\rho} \cos c, \quad H_c^\alpha \cos\varphi + H_s^\alpha \sin\varphi = \frac{r}{\rho} \cos b.$$

Si au lieu des corrections dM_0, $d\mu$, $d\varphi$, ds, dp, dq on introduit les nouvelles variables x, y, z, t, u, φ, liées aux valeurs précédantes par les corrélations:

$$x = (0.19803)\, dM_0, \quad z = (0.46432)\, d\varphi, \quad u = (0.16325)\, dp,$$
$$y = (4.22205)\, d\mu, \quad t = (0.16453)\, ds, \quad \varphi = (0.14137)\, dq,$$

et si l'on exprime les membres libres non pas en secondes de l'arc, mais en une unité dont le logarithme est 2.60109, on obtient de nouvelles équations conditionnelles, dont les coefficients seront plus uniformes.

Sans écrire ces nouvelles équations conditionnelles, nous donnerons directement les équations normales qui servent à calculer les valeurs x, y, z, t, u, φ.

Ces équations normales sont:

$+27.7194\,x$	$+4.1048\,y$	$-0.5603\,z$	$+29.5300\,t$	$+0.0076\,u$	$-0.0240\,\varphi$	$=-5.8188$
$+4.1048$	$+11.3726$	-0.3309	$+4.1120$	$+0.0038$	-0.0032	-4.4674
-0.5603	-0.3309	$+15.0007$	-0.5222	$+0.0032$	-0.0027	$+0.8105$
$+20.5300$	$+4.1120$	-0.5222	$+31.6408$	$+0.0080$	-0.0243	-5.9838
$+0.0076$	$+0.0038$	$+0.0032$	$+0.0080$	$+14.8362$	$+0.7785$	-0.1322
-0.0240	-0.0032	-0.0027	-0.0243	$+0.7785$	$+10.1297$	-0.2634

La solution des équations normales pour les inconnues et leurs poids a donné les valeurs suivantes:

$\log\varphi = 8.13800\,n$,	$\log p_\varphi = 1.28078$,	$\log z = 8.57327$,	$\log p_z = 1.17484$
$\log u = 7.90464\,n$,	$\log p_u = 1.17039$,	$\log y = 9.50083\,n$,	$\log p_y = 1.01661$
$\log t = 9.85268$,	$\log p_t = 9.24447$,	$\log x = 9.96431\,n$,	$\log p_x = 9.18399$

Cela va sans dire que pour la composition des équations normales et leur solution ont s'est servi de tous les contrôles possibles. Citons en un, à savoir $[ll\mathfrak{o}] = +1.0875$ et $[\delta\delta] = +1.0877$, où les lettres l et δ signifient, conformément, les membres libres et les erreurs résiduelles des équations conditionnelles.

En passant des inconnues x, y, z, t, u, v aux corrections dM_0, $d\mu$, $d\varphi$, ds, dp, dq et enfin en substituant les corrections des éléments di, $d\Omega$, $d\omega$ aux corrections ds, dp, dq, on obtient finalement pour l'époque initiale

$$dM_0 = -3'53''.01, \quad d\varphi = +5''.13, \quad d\Omega = -2'35''.09,$$
$$d\mu = -0''.00758, \quad di = -1''.24, \quad d\omega = +5'49''.67.$$

Donc les éléments définitifs seront:

T 1934 Mai 6.0, temps universel

$i = 1°35'59''.03$	équinoxe moyen 1925.0	$M_0 = 32°\ 4'3''.76$,	$\mu = 615''.06382$
$\Omega = 178\ 44\ 44\ .48$			
$\omega = 359\ 52\ 23\ .66$		$\varphi = 3\ 27\ 5\ .04$,	$\log a = 0.507391$

Enfin, les erreurs résiduelles des équations conditionnelles en parties du grade représentant les écarts des observations de la plus probable orbite de la planète Gerda, que j'ai obtenue, sont données dans la table suivante.

Temps moyen de Berlin		$\Delta\alpha \cos\delta$	$\Delta\delta$
1872 Août	29.0	+0°.043	+0°.012
1873 Octobre	26.0	+0 .032	+0 .010
1876 Avril	8.0	0 .000	+0 .002
1877 Juin	30.0	+0 .005	−0 .001
1878 Septembre	19.0	+0 .009	+0 .001
1885 Décembre	9.0	−0 .009	−0 .003
1893 Mars	16.0	−0 .018	+0 .008
1896 Novembre	5.0	−0 .009	−0 .006
1897 Décembre	25.0	−0 .010	+0 .001
1899 Mars	18.0	−0 .017	+0 .008
1901 Août	15.0	−0 .005	−0 .003
1902 Novembre	11.0	−0 .005	+0 .003
1904 Janvier	15.0	−0 .001	+0 .002
1905 Avril	7.0	−0 .008	+0 .004
1906 Juillet	8.0	−0 .008	0 .000
1907 Septembre	14.0	−0 .003	−0 .003
1908 Novembre	24.0	−0 .002	0 .000
1910 Février	5.0	−0 .058	+0 .003
1911 Mai	3.0	−0 .004	+0 .002
1913 Octobre	10.0	−0 .004	−0 .003

Temps moyen de Greenwich		$\Delta\alpha \cos\delta$	$\Delta\delta$
1916 Mars	16.0	+0°.001	−0°.003
1917 Mai	25.0	−0 .004	0 .000
1918 Août	1.0	−0 .004	−0 .004
1921 Janvier	19.0	+0 .013	+0 .002
1922 Mai	23.0	+0 .005	0 .000
1923 Mai	29.0	−0 .012	0 .000
1924 Novembre	29.0	−0 .009	−0 .005
Temps universel			
1927 Janvier	7.0	+0 .027	−0 .001
1928 Avril	13.0	+0 .013	−0 .004
1929 Juillet	6.0	−0 .014	−0 .003
1930 Septembre	19.0	−0 .008	−0 .001
1931 Novembre	12.0	+0 .006	+0 .001
1933 Février	25.0	+0 .055	−0 .023
1934 Mai	13.0	+0 .009	−0 .003

La table des erreurs résiduelles montre que la plus probable orbite de la planète Gerda, que j'ai obtenue, représente d'une manière si satisfaisante du point de vue pratique les observations faites pendant les 62 années, au cours desquelles elle fut observée, qu'on pourrait déjà poser la question de dresser les tables, par exemple, d'après la méthode de Backlund, pour déterminer les positions approximatives de la planète pour des périodes à venir considérables, avec une précision suffisante pour la pratique.

Manuscrit reçu
le 19.IV. 1937.

LITTÉRATURE CITÉE

[1] A. Ivanov et E. Martjanova-Ivanova, Bull. de l'Acad. des Sci. de l'URSS, 205—214 (1928). [2] A. A. Ivanov, C. R. Acad. Sci. URSS, I, No. 1, 11—16 (1935). [3] J. Bauschinger, Die Bahnbestimmung der Himmelskörper, Leipzig, 424—431 (1906).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV. № 6—7

MÉCANIQUE

FORMULES FONDAMENTALES POUR LE CALCUL DES ÉLÉMENTS DE LA TRAJECTOIRE DU CENTRE DE GRAVITÉ D'UN PROJECTILE

Par M. KURENSKII (KOURENSKY)

(Présenté par A. N. Krylov, m. de l'Académie, le 3. IV. 1937)

1. L'application du procédé proposé dans la note précédente pour la solution du problème sur le calcul des éléments de la trajectoire du centre de gravité d'un projectile, comportant, pour une formule à terme unique de la fonction de la résistance de l'air, $F(v) = av^n$, la transformation de la trajectoire en ponts, c'est-à-dire quand nous pouvons obtenir une approximation suffisante pour le calcul pratique au moyen de la formule de Taylor:

$$F(v) = 1 - kv + k^2v^2; \quad k = 0{,}00021,$$

nous parvenons au calcul des fonctions $v_0, v_1, v_2, \ldots, x_0, x_1, x_2, \ldots$ donnant la vitesse $v(t)$ et l'abscisse $x(t)$

$$v = v_0 + \lambda v_1 + \lambda^2 v_2 + \ldots; \quad x = x_0 + \lambda x_1 + \lambda^2 x_2 + \ldots,$$

où $\lambda = \frac{ak}{g}$ est un très petit paramètre. Tout les calculs ne sont nécessaires pour le calcul pratique, qu'à partir du quatrième terme de la série. Les fonctions v et y doivent vérifier le système d'équations:

$$\frac{dv}{dt}\cos\theta = -v\sin\theta + \lambda v^2(1 - kv - k^2v^2); \quad \frac{dx}{dt} = \frac{dy}{dt}\operatorname{tg}\theta = v,$$

de même qu'aux conditions initiales:

$$t = t_0, \quad v = V, \quad y = 0.$$

La distance $x(t)$ et la flèche du trajet $y(t)$ sont déterminées par les quadratures au moyen des équations

$$\frac{dx}{dt} = \frac{dy}{dt}\operatorname{tg}\theta, \quad \frac{dx}{d\theta} = \frac{v^2}{g\cos\theta}$$

et les conditions initiales

$$\theta = \theta_0; \quad x = 0; \quad t = 0.$$

2. En faisant varier les valeurs de l'exposant n et du coefficient a dans la formule $F(v) = av^n$ au passage à une autre loi de résistance, on peut obtenir une approximation suffisante pour le calcul pratique sans nombre de cas, en négligeant les termes comportant de très petits produits

$$k^2v^2; \quad k^3v^3; \quad k^4v^4; \ldots$$

En comparant les coefficients à puissances égales du paramètre λ, on obtiendra les équations suivantes, faciles à intégrer:

$$\left.\begin{aligned} v_0' \cos\theta = v_0 \sin\theta; \quad & y_0' + \frac{v_0^2}{g}\operatorname{tg}\theta = 0; \\ v_1' \cos\theta = v_1 \sin\theta + v_0^{n+1}(1 - ky_0); \quad & y_1' + \frac{2v_0 v_1}{g}\operatorname{tg}\theta = 0; \\ v_2' \cos\theta = v_2 \sin\theta + (n+1)\, v_0^n\, v_1; \quad & y_2' + \frac{v_1^2 + 2v_0 v_2}{g}\operatorname{tg}\theta = 0. \end{aligned}\right\} \qquad (1)$$

Nous avons:

$$\left.\begin{aligned} v_0 &= \frac{V\cos\theta_0}{\cos\theta}; \quad v_1 = -\frac{V^{n+1}\cos^{n+1}\theta_0}{\cos\theta}(\xi_n^0 - \xi_n) + \\ &+ \frac{kV^{n+3}\cos^{n+3}\theta_0}{2g\cos\theta}\left[\frac{\xi_n^0 - \xi_n}{\cos^2\theta_0} - (\xi_{n+2}^0 - \xi_{n+2})\right]; \\ v_2 &= \frac{(n+1)\,V^{2n+1}\cos^{2n+1}\theta_0}{2g\cos\theta}(\xi_n^0 - \xi_n)^2; \\ y_0 &= \frac{V^2}{2g}\left(1 - \frac{\cos^2\theta_0}{\cos^2\theta}\right); \quad y_1 = \frac{V^{n+2}\cos^{n+2}\theta_0}{g}\left[\frac{\xi_n^0 - \xi_n}{\cos^2\theta} - (\xi_{n+2}^0 - \xi_{n+2})\right] - \\ &- k\cdot\frac{V^{n+4}\cos^{n+4}\theta_0}{2g^2}\left\{\left[\frac{\xi_n^0 - \xi_n}{\cos^2\theta_0} - (\xi_{n+2} - \xi_{n+2})\right]\frac{1}{\cos^2\theta} - \right. \\ &\left. - \frac{\xi_{n+2} - \xi_{n+2}}{\cos^2\theta_0} + (\xi_{n+4}^0 - \xi_{n+4})\right\}, \\ y_2 &= \frac{(n+2)\,V^{2n+2}\cos^{2n+2}\theta_0}{g}\left\{(\xi_n - \xi_n)\left[\frac{\xi_n}{\cos^2\theta_0} - \right.\right. \\ &\left. - \frac{n+1}{2(n+2)}(\xi_n^0 + \xi_n)\right] - \frac{1}{n+2}\left(\frac{\xi_n^0\sin\theta_0}{\cos^{n+2}\theta_0} - \frac{\xi_n\sin\theta}{\cos^{n+2}\theta}\right) + \\ &+ \xi_n^0(\xi_{n+2}^0 - \xi_{n+2}) + \frac{1}{2(n+1)(n+2)}\left(\frac{1}{\cos^{2n+2}\theta_0} - \frac{1}{\cos^{2n+2}\theta}\right)\Bigg\}; \\ t_0 &= \frac{V\cos\theta_0}{g}(p_0 - p); \quad t_1 = \frac{V^{n+1}\cos^{n+1}\theta_0}{g}\left[(\xi_n^0 - \xi_n)\,p - \right. \\ &\left. - \frac{1}{n+1}\left(\frac{1}{\cos^{n+1}\theta_0} - \frac{1}{\cos^{n+1}\theta}\right)\right] + \\ &+ k\cdot\frac{V^{n+3}\cos^{n+3}\theta_0}{2g^2}\left[p\left(\xi_{n+2}^0 - \xi_{n+2} - \frac{\xi_n^0 - \xi_n}{\cos^2\theta_0}\right) - \frac{1}{n+3}\left(\frac{1}{\cos^{n+3}\theta_0} - \right.\right. \\ &\left.\left. - \frac{1}{\cos^{n+3}\theta}\right) + \frac{1}{(n+1)\cos^2\theta_0}\left(\frac{1}{\cos^{n+1}\theta_0} - \frac{1}{\cos^{n+1}\theta}\right)\right], \\ t_2 &= -\frac{(n+1)\,V^{2n+1}\cos^{2n+1}\theta_0}{2g}\left\{(\xi_n^0 - \xi_n)\left[\frac{2}{(n+1)\cos^{n+1}\theta_0} + \right.\right. \\ &\left. + p(\xi_n^0 - \xi_n)\right] - \frac{2}{n+1}(\xi_{2n+1}^0 - \xi_{2n+1})\Bigg\}; \\ x_0 &= t_0 V\cos\theta_0; \quad x_1 = t_1 2V\cos\theta_0; \\ x_2 &= t_2 2\cdot\frac{n+2}{n+1}\cdot V\cos\theta_0; \end{aligned}\right\} \qquad (2)$$

où

$$p = \operatorname{tg}\theta; \quad p_0 = \operatorname{tg}\theta_0;$$

$$\xi_m = \int_0^{\theta}\frac{d\theta}{\cos^{m+1}\theta}; \quad \xi_m^0 = \int_0^{\theta_0}\frac{d\theta}{\cos^{m+1}\theta}.$$

Les formules plus exactes pour le calcul seront obtenues, si l'on détermine les termes à deuxième puissance de k dans les expressions pour v_1, y_1, x_1, t_1, ainsi que les termes à première et deuxième puissances de k pour v_2, y_2, etc.

3. Il existe des tables à 5 chiffres de la fonction balistique $\xi_m(\theta)$ pour les angles θ à intervalle de 1° et pour les exposants

$$m = 0,\ 1,\ 1.55,\ 1.7,\ 2,\ 3,\ \ldots,\ 8.$$

Pour les formules ci-dessus, en négligeant les termes $k^2\lambda^2$, $k\lambda^2$ et ceux plus grands, pour le calcul de x et t, ces tables de la fonction $\xi_m(\theta)$ devront être prolongées jusqu'à $m = 11$, si l'on prend en considération l'exposant maximum $n = 5$. Si l'on retient les termes comportant le multiplicateur $k\lambda^2$, il faudra avoir, afin de pouvoir déterminer x_2, t_2, des valeurs de $\xi_m(\theta)$ pour $m = 2n + 3 = 13$. Il sera aussi nécessaire de réduire l'intervalle de θ jusqu'à 30′.

Les équations différentielles pour toutes les valeurs entières des puissances positives de k et λ s'intègrent en quadratures. Toutes les quadratures s'expriment rationnellement par fonctions trigonométriques et par $\xi_m(\theta)$ pour toutes les valeurs de m. Il existe des formules qui permettent de calculer ξ_m au moyen de valeurs à indices moindres, par exemple

$$\xi_m = \frac{\sin\theta}{m\cos^m\theta} + \frac{m-1}{m}\xi_{m+2}.$$

4. Pour le cas particulier de $\lambda = 0$, c'est-à-dire lorsqu'on n'opère qu'avec les premiers membres du développement, on obtient les formules de la théorie parabolique, appliquées assez souvent dans la pratique:

$$\bar{v} = v_0 = \frac{V\cos\theta_0}{\cos\theta};\quad \bar{y} = y_0 = \frac{V^2}{2g}\left(1 - \frac{\cos^2\theta_0}{\cos^2\theta}\right);$$

$$\bar{x} = x_0 = \frac{V^2\cos^2\theta_0}{g}(p_0 - p);\quad \bar{t} = t_0 = \frac{V\cos\theta_0}{g}(p_0 - p);$$

$$x_0 = t_0 V\cos\theta_0;\quad y_0 = t_0 V\sin\theta_0 - \frac{gt_0^2}{2} = x_0 p_0 - \frac{gx_0^2}{2V^2\cos^2\theta_0}.$$

Un autre cas particulier est celui où $k = 0$, ou c'est-à-dire lorsque la densité de l'air est constante. On omet alors dans les formules les membres avec k. De plus, les équations différentielles nous conduisent à la solution du problème d'apres Bernoulli:

$$\left.\begin{aligned} \bar{v} &= \frac{V\cos\theta_0}{\cos\theta}U(\theta_0,\ \lambda_g V^n,\ \theta);\quad \bar{y} = \frac{V^2\cos^2\theta_0}{g}\int_0^{\theta_0}\frac{U^2\operatorname{tg}\theta}{\cos^3\theta}\,d\theta;\\ \bar{x} &= \frac{V^2\cos^2\theta_0}{g}\int_0^{\theta_0}\frac{U^2 d\theta}{\cos^2\theta};\quad \bar{t} = \frac{V\cos\theta_0}{g}\int_0^{\theta_0}\frac{U d\theta}{\cos^2\theta}, \end{aligned}\right\} \qquad (3)$$

où

$$U = [1 + \lambda n V^n\cos^n\theta_0(\xi_n^0 - \xi_n)]^{-\frac{1}{n}}.$$

Par le fait que le système des équations différentielles ne doit comporter qu'une solution unique, nous sommes conduits à la formule, permettant de déterminer les premiers termes des expressions pour v_1, v_2, v_3, ... qui ne comportent point k, au moyen de la formule du binôme:

$$[1 + \lambda n V^n\cos^n\theta_0(\xi_n^0 - \xi_n)]^{-\frac{1}{n}} = 1 + \frac{\lambda v_1\cos\theta}{V\cos\theta_0} + \lambda^2\frac{v^2\cos\theta}{V\cos\theta_0} + \ldots =$$

$$= 1 - \lambda V^n\cos^n\theta_0(\xi_n^0 - \xi_n) + \lambda^2(n+1)V^{2n}\cos^{2n}\theta_0\frac{(\xi_n^0 - \xi_n)^2}{2} - \ldots$$

Les résultats connus des conditions de valabilité de la formule du binôme nous amènent à conclure à la convergence des séries. obtenues par développement par les puissances du paramètre λ. Il ne faut pas oublier que du commencement jusqu'au sommet de la trajectoire $\xi_n^0 > \xi_n$ tandis qu'au point de chute il se peut que $\xi_n^0 < \xi_n$. Si on applique la formule max $(\xi_n^0 - \xi_n) = \xi_n^0$ en substituant pour λ. V. θ_0. ξ_n^0 leurs valeurs numériques et en calculant successivement les membres d'après la formule du binôme. on pourra facilement établir le numéro du membre $\lambda^n c_n$. qu'il faudra prendre pour le calcul afin d'obtenir le degré d'approximation voulu.

Par exemple. pour le tir d'un obusier de 120 mm. pour les données suivantes:

$$V_0 = 183 \text{ m/sec}; \quad \theta_0 = 40°; \quad d = 0.120 \text{ m}; \quad q = 22 \text{ kg};$$
$$i = 0.668; \quad \Pi = \Pi_0 = 1.206 \text{ kg/m}^3; \quad g = 9.81 \text{ m/sec}^2,$$

l'on aura:

$$\max \lambda \cdot 2V^2 \cos^2 \theta_0 (\xi_2^0 - \xi_2) \approx 0.197 < 0.2;$$

$$(1 + 0.2)^{-\frac{1}{2}} = 1 - 0.1 + 0.015 - 0.0025 + \ldots$$

D'autre part. ces formules permettent de calculer. assez facilement et avec l'approximation voulue. les fonctions v. y. x. t pour les exposants n. pour lesquels il n'y a pas dans la méthode de Bernoulli de tables des fonctions balistiques correspondantes ($n = 2. 3. 4$). sans avoir recours à des formules compliquées au moyen desquelles ces fonctions sont calculées pour $n = 2. 3. 4$.

5. Pour le cas de densité variable de l'air. l'on retiendra les deuxièmes puissances du paramètre k et l'on aura. au lieu de la 3-ème et de la 5-ème équations du système (1):

$$v_1' \cos\theta = v_1 \sin\theta + v_0^{n+1}(1 - ky_0 + k^2 y_0^2);$$
$$v_2' \cos\theta = v_2 \sin\theta + v_0^n \{(n+1) v_1 - k[y_1 v_0 +$$
$$+ (n+1) y_0 v_1] + k^2 [2y_1 v_0 + y_0 (n+1) v_1] y_0 \}.$$

Au lieu de (2) et (3). on obtiendra des formules plus exactes pour le calcul. si l'on additionne tous les termes comportant comme multiplicateurs toutes les puissances d'un seul paramètre λ. et si l'on retient les termes comportant les produits des premières puissances de λ et k. L'on aura ainsi:

$$v = \overline{v} + \alpha; \quad y = \overline{y} + \beta; \quad x = \overline{x} + \gamma; \quad t = \overline{t} + \delta.$$

Les corrections α. β. γ. δ pour les formules (3) ont la forme suivante:

$$\alpha = \lambda k \cdot \frac{V^{n+3} \cos^{n+3} \theta_0}{2g \cos\theta} \cdot (a_1 + a_2 k + a_3 \lambda);$$

$$a_1 = \frac{\xi_n^0 - \xi_n}{\cos^2 \theta_0} - (\xi_{n+2}^0 - \xi_{n+2}),$$

$$a_2 = -\frac{V^2}{2g}\left[\frac{\xi_n^0 - \xi_n}{\cos^2 \theta_0} - 2(\xi_{n+2}^0 - \xi_{n+2}) + \cos^2 \theta_0 (\xi_{n+4}^0 - \xi_{n+4})\right].$$

$$a_3 = -V^n \cos^n \theta_0 \left[(n+1)\left(\frac{\xi_n^0 - \xi_n}{\cos\theta_0}\right)^2 -\right.$$

$$- (n+3)(\xi_{n+2}^0 - \xi_{n+2})\left(\xi_n^0 - \frac{n-1}{n-3}\xi_n\right) + 2 \cdot \frac{n+1}{n-2}(\xi_n^{02} - \xi_n^2) +$$

$$\left. + \frac{4}{n-2}\left(\frac{\xi_n^0 \sin\theta_0}{\cos^{n-2}\theta_0} - \frac{\xi_n \sin\theta}{\cos^{n-2}\theta}\right) - \frac{2}{n-1}\frac{1}{n+2}\left(\frac{1}{\cos^{2n-2}\theta_0} - \frac{1}{\cos^{2n+2}\theta}\right)\right];$$

$$\beta = \lambda k \cdot \frac{V^{n+4} \cos^{n+4} \theta_0}{2g^2} \cdot (b_1 + b_2 k + b_3 \lambda);$$

$$b_1 = -\left[\frac{\xi_n^0 - \xi_n}{\cos^2\theta} - (\xi_{n+2}^0 - \xi_{n+2})\left(1 + \frac{\cos^2\theta_0}{\cos^2\theta}\right) + \cos^2\theta_0\,(\xi_{n+4}^0 - \xi_{n+4})\right],$$

$$b_2 = \frac{V^2}{2g}\left[\frac{\xi_n^0 - \xi_n}{\cos^2\theta} - (\xi_{n+2}^0 - \xi_{n+2})\left(1 + 2\frac{\cos^2\theta_0}{\cos^2\theta}\right) + \right.$$

$$\left. + (\xi_{n+4}^0 - \xi_{n+4})\left(2 + \frac{\cos^2\theta_0}{\cos^2\theta}\right)\cos^2\theta_0 - (\xi_{n+6}^0 - \xi_{n+6})\cos^4\theta_0\right]$$

$$b_3 = V^n \cos^n\theta_0 \cdot \text{fonct}\,(\xi_n^0, \xi\ .\ .\ .\ .\ , \xi_{2n+4}^0, \xi_{2n+4});$$

$$\delta = \lambda k \cdot \frac{V^{n+3}\cos^{n+3}\theta_0}{2g^2} \cdot\ d_1 + d_2 k + d_3\lambda);$$

$$d_1 = -\left\{\left[\frac{\xi_n^0 - \xi_n}{\cos^2\theta_0} - (\xi_{n+2}^0 - \xi_{n+2})\right]p + \right.$$

$$\left. + \frac{1}{n+3}\left(\frac{1}{\cos^{n+3}\theta_0} - \frac{1}{\cos^{n+3}\theta}\right) - \frac{1}{(n+1)\cos^2\theta_0}\left(\frac{1}{\cos^{n+1}\theta_0} - \frac{1}{\cos^{n+1}\theta}\right)\right\},$$

$$d_2 = \frac{V^2}{2g}\left\{\left[\frac{\xi_n^0 - \xi_n}{\cos^2\theta_0} - 2\,(\xi_{n+2}^0 - \xi_{n+2}) + (\xi_{n+4}^0 - \xi_{n+4})\cos^2\theta_0\right]p - \right.$$

$$- \frac{1}{(n+1)\cos^2\theta_0}\left(\frac{1}{\cos^{n+1}\theta_0} - \frac{1}{\cos^{n+1}\theta}\right) +$$

$$\left. + \frac{2}{n+3}\left(\frac{1}{\cos^{n+3}\theta_0} - \frac{1}{\cos^{n+3}\theta}\right) - \frac{\cos^2\theta_0}{n+5}\left(\frac{1}{\cos^{n+5}\theta_0} - \frac{1}{\cos^{n+5}\theta}\right)\right\},$$

$$d_3 = V^n \cos^n\theta_0 \cdot \text{fonct}\,(\xi_n^0, \xi_n, \ \ldots\ , \xi_{2n+3}^0, \xi_{2n+3});$$

$$\gamma = \lambda k \cdot \frac{V^{n+4}\cos^{n+4}\theta_0}{g^2} \cdot (c_1 + c_2 k + c_3\lambda).$$

6. Les expressions de b_3, c_3, d_3 exigent: 1°, un calcul au moyen de formules compliquées et 2°, la présence des tables de la fonction $\xi_m(\theta)$ pour $m = 2n + 4$. L'expression pour a_3 détermine l'erreur dans le calcul de la vitesse v, pour laquelle on prendra la formule $\alpha = \lambda k \cdot A\,(a_1 + k a_2)$, en posant $\lambda b_3 \approx 0$, $\lambda c_3 \approx 0$, $\lambda d_3 \approx 0$. L'on aura alors au lieu de γ et δ:

$$\gamma_1 = \lambda k \cdot C\,(c_1 + k c_2); \quad \delta = \lambda k \cdot D\,(d_1 + k d_2); \quad \gamma_1 = \delta_1 \cdot 2V\cos\theta_0.$$

Par exemple, pour le tir d'un mortir de 9′, quand

$$n = 3; \quad V = 315.5\ \text{m/sec}; \quad \theta_0 = 43°30'; \quad d = 0.2286\ \text{m};$$
$$q = 110.9\ \text{kg}; \quad i = 0.896; \quad g = 9.81\ \text{m/sec}^2,$$

l'on aura

$$\log\lambda = \overline{8}.3466; \quad \log k = \overline{4}.0414.$$

Pour le calcul de la vitesse immédiatement au sommet, on aura la correction α pour la variation de la vitesse, liée avec la variation de densité de l'air du commencement jusqu'au sommet de la trajectoire:

$$a_1 + a_2 k + a_3\lambda = 0.06; \quad \alpha = 1.09\ \text{m/sec}.$$

Institut des Mathématiques et de Mécanique
de l'Université de Leningrad.

Manuscrit reçu
le 5.IV.1937.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1 937. Volume XV, № 6—7

TECHNICAL PHYSICS

A METHOD OF ACOUSTICAL DESIGN OF ROOMS EQUIPPED WITH LOUD SPEAKERS

By V. FURDUJEV

(Communicated by V. Mitkevitč, Member of the Academy, 2. IV. 1937)

The sound field produced by a loud speaker in a close room may be defined by the density ε of sound energy at any point of the field. It is convenient to represent this quantity as the sum of two components: ε_D, the density of sound energy in a direct wave reaching the point under consideration without reflections, and ε_R, the density of energy transported by all the reflected waves. These components are defined by the expressions:

$$\varepsilon_D = \frac{p^2_{\varphi_1}}{\rho c^2} = \frac{p_0^2}{\rho c^2} f^2(\varphi_1), \tag{1}$$

$$\varepsilon_R = \frac{4L}{c\alpha S}(1-\alpha), \tag{2}$$

where p_{φ_1} is the r. m. s. sound pressure at a point lying at the angle φ_1 off the normal loud speaker axis, p_0—sound pressure at the equidistant point on this axis, $f(\varphi)$—radiation distribution characteristic, which is assumed to be symmetrical with respect to the loud speaker axis, L—acoustic power output, ρ—density of air, c—velocity of sound, α—average absorption coefficient, S—area of absorbing material.

Neglecting the reaction of room upon the source of sound, we can write for acoustic power radiated within the solid angle 2π, the expression:

$$L = 2\pi D_1^2 \frac{p_0^2}{\rho c} \int_0^{\frac{\pi}{2}} f^2(\varphi) \sin\varphi \, d\varphi \tag{3}$$

(D_1 is the distance where the axial pressure is p_0). Combining (1), (2) and (3), we obtain the ratio of reflected to direct sound which defines the contribution of the diffuse sound field to the sensation of a listener:

$$R = \frac{\varepsilon_R}{\varepsilon_D} = \frac{8\pi}{S}\left[\frac{D_1}{f(\varphi_1)}\right]^2 \frac{1-\alpha}{\alpha} \int_0^{\frac{\pi}{2}} f^2(\varphi) \sin\varphi \, d\varphi. \tag{4}$$

The acoustical design of rooms equipped with loud speakers must be based upon formula (4). The following requirements should be fulfilled:

1) The acoustic ratio should not be too great, since a distinct reproduction of sound is required. 2) The acoustic ratio should be constant through

the reproduced frequency range. 3) The acoustic ratio should be constant over the whole seating area. The conditions of possibility to satisfy these requirements can be derived from formula (4).

In order to decrease R it is convenient to decrease the reverberation time, i. e. to increase the average absorption coefficient α. Formula (4) shows another possibility of decreasing the acoustic ratio, namely, the use of a directionall loud speaker for which

$$\int_0^{\frac{\pi}{2}} f^2(\varphi) \sin\varphi \, d\varphi < 1.$$

In order to reach a constant acoustic ratio with varying frequency, we are obliged to renounce the use of flat baffle loud speakers having a radiator which is small as compared with wave length at either low or medium frequencies. If we represent such a loud speaker as a piston diaphragm radiating into semi-infinite space, the radiation distribution characteristic may be written thus: $f(\varphi) = \frac{2I_1(ka \sin\varphi)}{ka \sin\varphi}$, where I_1 is the Bessel function of the first order, $k = \frac{2\pi}{x}$ and a is the radius of the diaphragm; thus, the frequency characteristic of R (if we neglect the frequency curve of absorption) is defined by the integral:

$$\int_0^{\frac{\pi}{2}} \left[\frac{2I_1(ka \sin\varphi)}{ka \sin\varphi}\right]^2 \sin\varphi \, d\varphi = \frac{2}{k^2a^2}\left[1 - \frac{I_1(2ka)}{ka}\right]. \qquad (5)$$

The expression (5) shows that, under these conditions, the quantity R decreases with increasing frequency, at first slowly and further very rapidly. Thus, the fulfillment of the second requirement becomes impossible when using a loud speaker whose directional properties vary rather rapidly as the frequency increases.

Using a horn loud speaker, it is possible to reach a constant acoustic ratio through a certain range of frequencies with approximation sufficient for technical purposes, since the form of the radiation distribution characteristic shows in this case only slight variation with frequency. At relatively high frequencies, the main portion of sound energy is radiated from the diaphragm in the form of a beam diverging slower than the walls of the horn; thus, as the frequency increases, the «radiating» section of the horn becomes smaller.

Now let us consider the condition of constancy of acoustic ratio over the seating area. We have seen that the coordinates of a listener enter into formula (4) in form of the ratio $\frac{D}{f(\varphi)}$. Thus, it is always possible to find such a disposition of listeners with respect to the loud speaker axis where the distance D would vary from point to point following the same law as the variation of the function $f(\varphi)$, the ratio $\frac{D}{f(\varphi)}$ remaining now constant. Thus, the constancy of acoustic ratio can be reached when the disposition of listeners along and across the loud speaker axis reproduces the form of the radiation distribution characteristic. Inversely, for a given disposition of listeners, it is always possible to build such a radiation distribution characteristic for which the function $f(\varphi)$ varies from point to point following the same law as the variation of distance between the sound source and a listener.

Research Institute of Motion Pictures and Photography.
Acoustical Committee of the Academy of Sciences of the USSR.
Received
2.IV.1937.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

TECHNICAL PHYSICS

EXPERIMENTAL INVESTIGATIONS OF REVERBERATION OPTIMUM FOR DIFFERENT FREQUENCIES

By S. LIFSHITZ

(Communicated by V. Mitkevič, Member of the Academy, 2. IV. 1937)

The question of reverberation optimum for different frequencies has interested a considerable number of investigators both in the Soviet Union and abroad during the past seven years. The author presented the complete theory of the question in 1934 (1) and 1936 (2,3). The curves of MacNair (4) and Knudsen (5) and of other investigators can be regarded as individual cases of the general theory of the author.

All of the above-mentioned investigators agree that the reverberation optimum for different frequencies is curvilinear, and that the optimum for low frequencies and for high frequencies is longer than that for medium ones. According to the author's theory, this curvilinearity of reverberation optimum for different frequencies indicates that we receive artistic impression of musical performances only when the apparent duration of tone is the same for all frequencies and is equivalent to 41 units of apparent duration. In this way, the basic tone and all overtones simultaneously have one and the same apparent duration of 41 units. In planning and constructing concert halls, theatres, radio studios and radio theatres in the Soviet Union, England, the United States and Germany, one is in most cases guided by the curvilinear characteristic of frequency optimum.

The conclusions Békésy arrives at, on the basis of his experimental work, are quite opposite to the conditions indicated above. In his works published in 1933, 1934 (6) and 1936 (7), Békésy used as sound absorbers special material (taut linen lined with cotton), which made it possible to change the curvature of the optimum for different frequencies and also to obtain a rectilinear characteristic. On the basis of his experimental work, Békésy comes to the following conclusions:

1. The best conditions for musical performance are obtained when the optimum reverberation for different frequencies in the hall is rectilinear, that is, the reverberation for all frequencies is one and the same.

2. The reverberation optimum for the average frequency (512 vib./sec.) obtained during rectilinear absorption is lower than the optimum under usual conditions, viz. 0.7—0.8 sec. instead of 1.06 sec. for small halls.

In spite of the many experiments carried out under good conditions, Békésy overlooked one vital mistake in method, the existence of which chan-

ges the value of the results obtained. Békésy found the optimum by changing the absorption in rooms equipped with his unusual absorbers. He was unable directly to compare the results with the optimum in halls having a curvilinear character. One can therefore presume that the optimum found by Békésy is a secondary optimum for unusual absorbers, and that it might be inferior to the usual optimal conditions with curvilinear absorption.

In order to verify Békésy's conclusions and at the same time avoid the mistake indicated above, we worked out the following method for experiments. In the Laboratory of Architectural Acoustics of the Moscow Architectural Institute two similar, adjacent rooms, each 100 m³ in volume, were equipped with absorbers. Room No 1 was furnished with Békésy absorbers. In this room it was possible to obtain a reverberation optimum that was rectilinear. Room No 2 was equipped with an absorber of a soft material. The MacNair-Lifshitz curvilinear reverberation optimum could be obtained in Room No 2. The same performer executed the same selection in each of the rooms in the presence of musical experts, who thus were able to judge of the quality of musical performance under different conditions.

In our experiments we paid special attention to the accuracy of measurement of the time of reverberation at various frequencies. The measurements

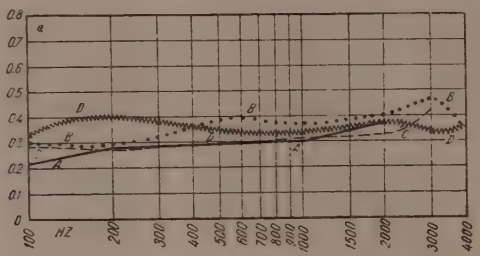


Fig. 1.

were made simultaneously by three methods: auditory (measurements were directed by S. Alexejev); the point method* (where the reverberation was determined by points; this was directed by Y. S. Schneider), and the thermocouple method (directed by S. Skrebkov). Good coincidence for the values obtained by these methods was achieved. An illustration of such coincidences is given in Fig. 1, showing the measurements of absorption coefficients of a sound-absorbing material prepared in the laboratory. In Fig. 1, line *A* represents the data obtained by the auditory method, line *B*, the data obtained by the point method, line *C*, the data obtained by the thermocouple method, and line *D*, the measurement of the same material by the point method in another laboratory (Physics Institute of the Academy of Sciences).

The first series of experiments consisted in obtaining a rectilinear optimum according to Békésy in Room No 1 and a curvilinear optimum according to MacNair-Lifshitz in Room No 2. At 512 vib./sec., the optimum was equal to 0.75 sec. in Room No 1, and to 1.03 sec. in Room No 2. The results of the measurements for all frequencies are shown in Fig. 2. The upper curve, *ML*, is the musical optimum according to MacNair-Lifshitz obtained in Room No 2, the lower band, *N*, is Békésy's optimum in Room No 1, this band showing the dispersion of individual points of the measurements.

---

* Similar to Knudsen's method involving the use of a relay-controlled meter ([8]).

Ten musical authorities were invited as experts, and they were present at the experiments at different times. These experts were: Composer Sergei Prokofjev; Prof. H. Neuhaus, director of the Moscow State Conservatory; Prof. Orlov, conductor; Prof. E. Braude, Honoured Art Worker; Composer Chemberdži; Conductor V. Kubatskij, Honoured Artist; Prof. N. Garbuzov, director of the Scientific Research Music Institute; E. Vitaček, Honoured Art Worker; P. Kozlov, head of the department of theory at the Gnesin Musical Technicum; and S. Skrebkov, lecturer at the Moscow State Conservatory.

A bass, a soprano, a tenor, a piano, a violin and a cello performed. When one heard the musical rendition in Room No 1 (Békésy), the performance seemed sufficiently artistic, but this impression changed immediately when

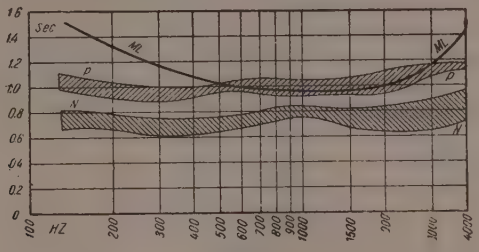


Fig. 2.

one went into Room No 2, with its curvilinear optimum. The performance in the second room sounded much more colourful, richer, more beautiful in timbre than that in the first. All of the 10 experts were unanimous in this opinion.

We quote excerpts of some conclusions reached by them:

S. Prokofjev: «Before listening to the performance I thought it would be difficult to distinguish which was better. But there was no difficulty, for everything in MacNair-Lifshitz Room No 2 sounded richer, rounder, more beautiful in timbre. The sound which in the Békésy Room was bare, acquired overtones of great beauty in Room No 2».

Prof. Orlov: «The performance in Room No 2 sounded much more colourful. The timbre was very soft and beautiful».

Prof. Neuhaus: «The music sounded better, fuller, more colourful in Room No 2. The sound seemed warmer and richer».

And so on.

These experiments show beyond doubt that the optimum obtained by Békésy is a secondary optimum, which gives the best sound for rectilinear characteristics. However, it is inferior to the optimum conditions which may be obtained with a curvilinear optimum.

The next series of experiments consisted in checking Békésy's statement that the optimum for 512 vib./sec. was about 0.75 sec. By changing the absorption in Room No 1, it was possible to get a rectilinear absorption at different magnitudes of reverberation for 512 vib./sec., and to listen to the musical performance at various magnitudes. Experiments have shown that in this case Békésy is right, and that the optimum for rectilinear absorption is below the generally accepted one.

Thus, for example, when the reverberation in the Békésy room was determined under conditions expressed by band *P* in Fig. 2, the musical performance was not so good as under the conditions indicated by band *N* in Fig. 2.

A comparison was made of the musical performance in Room No 1 under conditions indicated by band P with that in Room No 2 under conditions indicated by curve ML. The performance in Room No 1 suffered still more as compared with that in Room No 2. A new phenomenon was revealed this time: under the conditions expressed by band P, Room No 1 seemed more hollow than Room No 2 under the conditions expressed by the curve ML, although the reverberation in Room No 1 was for all frequencies, either the same or less than the reverberation in Room No 2. But that requires further investigations.

Conclusions

1. The rectilinear reverberation optimum provides less satisfactory conditions for musical performance than the usual curvilinear optimum.

2. The rectilinear reverberation optimum is below the usual. This very likely results from the fact that the room with a rectilinear optimum, when the same reverberation of 512 vib./sec. is used, sounds more hollow than the one with a curvilinear optimum.

Institute of Architecture.
Moscow.

Received
2. IV. 1937.

REFERENCES

[1] С. Лифшиц, ЖТФ, IV, вып. 9 (1934). [2] S. Lifshitz, Journal of the Acoustical Society of America, January (1936). [3] С. Лифшиц, ЖТФ, VI, вып. 12 (1936). [4] W. MacNair, J. A. S. A. (1930). [5] Vern Knudsen, J. A. S. A., April (1931). [6] Békésy, A. d. Physik (5), **16** (1933); ENT, No 11 (1934). [7] Békésy, J. T. Ph., **12** (1936). [8] Knudsen, J. A. S. A., **3**, 131 (1931).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

GEOPHYSICS

THE PRINCIPLES OF THE MONSOON THEORY

. SHOULEJKIN, Corresponding Member of the Academy

The monsoon air currents arising in winter ower the warm sea and the cold continent can be determined analytically in relation to the circular sea (surrounded by the large continent) or to the circular continent (encircled with the ocean). In the cylindrical system of co-ordinates, with the axis going through the centre of the sea, the radial component of the air velocity is determined by the equation:

$$u = U \cdot f(z) \tag{1}$$

where U is a function only of the radius-vector r:

$$U = -m \cdot r^{\nu} \cdot e^{-r}. \tag{2}$$

The radius r, as well as the height z over the sea level is expressed in the abstracted units: let us suppose that the unit-segment ρ is equal to the $(\nu+1)$th part of the radius of the sea. In the coastal zone, U is connected in a very simple manner with the pressure gradient $\frac{\partial p}{\partial r}$, the density of the air δ and the projection $\bar{\omega}$ of the angular velocity ω of the Earth rotation on the vertical line at the corresponding point of the globe:

$$U = \frac{\frac{\partial p}{\partial r}}{2\delta\bar{\omega}\rho}. \tag{3}$$

The function $f(z)$ only from the height over the sea level will be exactly determined in a next paper(2). At present, let us suppose schematically that:

$$f(z) = \sin az. \tag{4}$$

The constante a is connected with the so-called depth of friction D by means of relation: $a = \frac{\pi\rho}{D}$. The fundamental quantity D, according to Ekman (1), is represented by the formula:

$$D = \pi\sqrt{\frac{\mu}{\delta \cdot \bar{\omega}}},$$

where μ is the coefficient of the turbulent viscosity of the air. The vertical component of the air velocity is determined by the equation:

$$w = \frac{m}{a}[(\nu + 1) r^{\nu-1} - r^{\nu}] e^{-r} (1 - \cos az). \tag{5}$$

The tangential components v are of no importance in the investigation of the case of circular shape of the shore line: all the derivatives to azimuth amount to nought. According to (1), (2), (4) and (5), it is very easy to find the equation of the stream-lines (more correct — their projections on the vertical plane):

$$z = \frac{D}{\pi\rho} \arccos\left(1 - C\frac{e^{r}}{r^{\nu+1}}\right), \tag{6}$$

C — being a parameter.

The climatological gradient of atmospheric pressure is connected to that of the mean temperature of active stratum, by means of relation:

$$\operatorname{grad} p = -\Pi \operatorname{grad} \bar{\vartheta} \tag{7}$$

where:

$$\Pi = \frac{p_0}{T_0} \cdot \frac{gH}{RT_0 - gH}. \tag{8}$$

Here H denote the altitude of the active monsoon stratum, g — the acceleration in the field of gravity, R — the well known gas constant, T_0 — any absolute temperature and p_0 — any pressure, supposed as normal ones.

The mean temperature $\bar{\vartheta}$ of the active air stratum varies by increasing of the distance from the centre of the sea, according the law expressed in the equation:

$$\bar{\vartheta} = \theta\left[1 - \frac{(\nu, r)!}{\nu!}\right] + \vartheta_{\infty}. \tag{9}$$

Here ϑ_{∞} denote the temperature of the most cold point of the continent, θ — the temperature difference between the most warm point of sea and the most cold one of the continent.

The maximum value of the radial velocity $U_{\max}$ proves to be connected to the quantity Π [see (8)] and to θ:

$$U_{\max} = \frac{\left(\frac{\nu}{e}\right)^{\nu}}{2\nu!} \cdot \frac{\Pi\theta}{\delta\omega\rho}. \tag{10}$$

Thus, the value θ may be considered as something similar to any difference of potentials stipulating the motion of air masses as well as the heat flows.

The vertical distribution of the air temperature, in the coastal zone may be expressed by means of an approximate relation:

$$\vartheta = \frac{\beta}{a}\left(z - \frac{1}{a}\sin az - 1\right) + \bar{\vartheta}. \tag{11}$$

More explicite relation will be found immediately after correction of the kind of the function $f(z)$ in the equation (1). In the relation (11) the function U enters in hidden form. For:

$$\frac{\beta}{a} = \frac{2}{\pi} \cdot \frac{\delta\omega D}{K'\Pi} \cdot U^2. \tag{12}$$

Here K' denote a value, connected with the turbulent thermal conductivity k, unit-segment ρ, specific heat c and density of the air δ:

$$K' = \frac{\rho \cdot k}{c \cdot \delta}.$$

The amount of heat Q transferred by the air currents in a time unit inside the sector with the inside angle ψ is expressed by the equation:

$$Q = \frac{3}{\pi^2} \cdot \frac{c\bar{\omega}\delta^2}{K'\Pi} \cdot D^2\rho^2\psi \cdot U^3. \tag{13}$$

The amount of heat given away to the atmosphere by the sea or taken from the atmosphere by the continent (corresponding to different values of r) is expressed by the relation:

$$dq = \frac{9}{\pi^2} \cdot \frac{c\bar{\omega}\delta^2}{K'\Pi} \cdot D^2\rho^2\psi \cdot U^2 \cdot \frac{\partial U}{\partial r} \cdot dr, \tag{14}$$

this amount of heat referring to the segment of a ring-like zone dr-wide, limited on two other extremities by the radii forming as previously the angle ψ.

The condition according to which the amount of heat given away by the sea must exactly equal to that received by the continent, is satisfied automatically (if the loss of the heat in the interplanetary space is eliminated from our problem). For, departing from (14), this condition can be formulated very easy:

$$\int_0^\infty d(U^3) = 0. \tag{15}$$

But this condition is satisfied by every kind of function U where becomes equal to nought both in the central point of the sea and in infinitely remote ones.

The formula (11) being also in very good concordance with the results of the immediate recording of temperature in different altitudes, at the polar coastal meteorological stations (after elimination of the effect of the loss of heat in the interplanetary space). Any discrepances between they can serve to determination of corrections to the formula (4).

The velocity U calculated on the basis of the above theory according to the climatic isobars, proves to be rather similar to that observed in natural conditions.

Judging by theoretical considerations one could await a complete similarity between the charts of climatic isobars and those of the temperature isanomalic lines (2) for the corresponding winter months. The comparison of these two groups of charts shows indeed a likeness even in details and thus once more corroborates the theory.

On this two groups of charts it is easy to calculate two curves representing: first the annual variations of the pressure gradient and second — the annual variations of the gradient of temperature anomaly. Both the curves prove to be of the same type and displaced for 1.5 month, i. e. one eight of the period. This ought to be expected owing to the instationarity of processes in natural conditions.

The comparison of both this curves allows us to calculate by means of equation, similar to (7), the important value Π found in the theory.

After its calculation it is easy to calculate according to (8), i. e. according to

$$H = \frac{\Pi R T_0^2}{g(p_0 + \Pi T_0)}, \tag{16}$$

the actual altitude of the active monsoon stratum.

The altitude calculated proves to be of the same order (2 km) as the determined in natural conditions.

Being substituted in the relation (10), the same numerical value of Π allows us to determine the order of the anticipated maximum radial velocity U_{max}. As concerning to the important value θ — its numeral value is determined immediately from the charts of temperature isanomalic lines.

The calculated value proves to be of the same order as the velocity, measured in natural conditions, if for the exponent ν in the equation (2) is substituted a numeral value $\nu = 2$.

The amount of heat transferred by air currents over one linear centimetre of the coast line from the sea to the continent, during one winter month, can be calculated according the formula:

$$\Phi_2 = 2.6 \cdot 10^6 \cdot \eta \cdot c\delta HU \cdot \Delta\vartheta \tag{17}$$

where $\Delta\vartheta$ denote the temperature difference between the opper and under boundary of the active monsoon stratum (calculated after the elimination of the loss of heat into interplanetary space). The numeral coefficient η be proximate to 0.5.

The calculation by means the (17) in relation to the North Polar Bassin gives the character of value Φ_2 (10^{12} small calory) proved to be similar as was found by the calculation of the heat flows according to the method of turbulent heat conduction [2].

Department of Theoretical Geophysics.
Institute of Geography of the Academy of Sciences of the USSR.

Received
5. V. 1937.

REFERENCES

[1] V. W. Ekman, Ark. Mat. Astr. of Phys., 2, № 11 (1905). [2] W. W. Shoulejkin, Bull. Ac. Sci. USSR, № 8—9, 997 (1935) (Série Phys.-Math.).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

PHYSICS

A FLUOROMETER BASED ON THE EFFECT OF DEBYE AND SEARS

By L. A. TUMERMANN and V. ŠYMANOVSKIJ

(Communicated by S. I. Wawilow, Member of the Academy, 20. IV. 1937)

The existing fluorometers (1) which are all based on the principle of modulation of a light beam in a Kerr-cell filled with nitrobenzene have many defects. The first one is the relatively small intensity of the modulated light beam. The second is, that the measurements are limited only to the visible part of the spectrum, the ultra-violet rays being absorbed by the nitrobenzene. Besides experiments with Kerr-cells are difficult because of the heating up of the nitrobenzene and of the dependence of the Kerr constant from the wave length of light.

In order to avoid these defects another method of high frequency light modulation must be used instead of the Kerr effect. Trials of this kind were made by Duschinsky (2). He used the effect of the rotation of the plane of polarization in quartz as a function of the impressed voltage for modulation purposes. This idea was dropped, however, because of too small intensities of the modulated light beam. A similar set up seems to have been developed with success lately by Brüninghaus (3) but details of his investigations especially with regard to the question of intensities are still lacking.

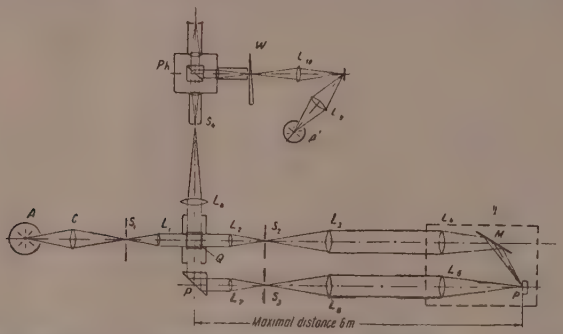
Mandelstamm, Landsberg and Papalexy and simultaneously with them Carolus proposed (1934) to use for modulation of a light beam effect of Debye and Sears namely the diffraction of light by ultra-sound waves. Recording to them such a modulation method could also be used for measuring the velocity of light. That such a modulation was experimentally possible at high frequency was shown by Becker, Hanle and Maerks (4) in a paper published during the time of our own investigations. Later a paper was published by Maerks (5) in which he gives, using a somewhat different set up as ours, an experimental curve corresponding to the well known «mirror curves» in fluorometric measurements. Maerks proposed such a set up as a new fluorometer. No measurements of any durations of luminescence were published with this method, however, up to the present time.

Independently of Maerks we applied a similar method of light modulation and developed a fluorometer based on the effect of Debye and Sears with a general set up as shown in the figure.

A light beam from an Osram point light lamp A (4A) passes through a slit s_1, a lens L_1 and quartz parallelepiped ($2\times3\times4$ cm). The quartz is pla-

ced between two brass electrodes which are connected with a short wave oscillator.

The oscillator can be very accurately tuned by means of a variable small capacity condenser so that the emmited frequency corresponds exactly to a mechanical overtone of the quartz. In this case standing ultra-sound waves will be developed in it and a diffraction pattern of the slit s_1 can be obtained at s_2. The energy in this diffraction pattern will fluctuate back and forth from the zero order to higher orders. The slit s_2 cuts off the higher diffraction orders and the zero order passes through lens L_3 to a small carriage which



A—point light lamp; A'—a comparing lamp; C—condenser; L_1—L_{10}—lenses; s_1—s_4—interstice; Q—quartz parallelepiped; M—mirror; P— prism of the total inner reflection; Ph—photometer with a cotter W. The section surrounded by dotted lines is placed on a moveable carriage.

is moveable on rails. This light beam is therefore modulated with a frequency 2ν if ν is the oscillators frequency.

The light beam is then either diffused in P or it excites there fluorescent radiation. The diffused light or the fluorescent radiation goes then through the lenses L_5, L_6, L_7 and slit s_3 to the quartz. There it is again diffracted. The higher diffraction orders are again screaned out the slit s_4 and only the zero order passes through it to a photometer Ph where its intensity is measured in an usual way (Lummer-Brodhun cube and wedge). The standard light was always of the same colour as the measured one.

It is evident that the measured intensity of the zero order after a second modulation will be depending upon the difference of the phases of the ultrasound standing waves at the time of the first and of the second passage of the light beam through the quartz, that is upon the distance of the carriage from it. For diffused light the minimum intensity must be at a distance equal to $\frac{\lambda}{8}$ (λ—emmited wave length of the oscillator). For fluorescent radiation this minimum will be nearer to the quartz correspondingly to the additional phase retardation caused by the time of decay τ. Having therefore the intensity curves for diffused light and for fluorescence as a function of a distance of the carriage from the quartz we can easily in a way described already by one of us ([6]) determine the average duration of luminescence from the displacements X of the minimas of both curves. If the frequency ν is sufficiently small with respect to τ we have then:

$$\tau = \frac{2X}{c}$$

when c *is* the velocity of light.

Preliminary measurements for a fluoresceine solution in water (concentration $\approx 10^{-4}$ g/cm^3) gave $\tau = 4.7 \cdot 10^{-9}$ sec. what is in good accord with former measurements (1).

In order to have an approximate idea of the accuracy of our measurements we determined the minimum for the diffused light curve and got it at a distance of circa 300 cm from the quartz which corresponds to an oscillator wave length $\lambda = 24$ m. The wave lenght measured with a wavemeter was $\lambda = 24.4$ m. We could therefore measure the velocity of light with an accuracy of circa 2%.

It must be noted that we tried, without much success, the method described by Maerks, that is the use of the diffraction by ultra-sound standing waves in a fluid rather than in quartz. In order to get namely sharp minima of both curves it is absolutely necessary to have a high degree of modulation of the light beam. This is again only possible if we gave a pure system of standing ultra-sound waves. At the frequencies used where the ultra-sound waves are of an order of hundreths of millimeter it is technically very hard to get pure standing waves in a fluid and therefore a high degree of modulation. Maerks also mentiones in his paper that he got a comparatively low degree of modulation of the zero order. Using ultra-sound standing waves in a quartz parallelepiped all these obstacles are surmounted and a very high degree of modulation of the zero order (over 80%) could be obtained.

The above described method has great possibilities for fluorometric measurements of the time of decay of fluorescence as of resonance radiation because:

1. It gives much higher intensities of the modulated light than the Kerr effect method. Even very faint fluorescence can be measured with a simple point light lamp excitation whereas formerly a 20 A carbon arc was hardly sufficient.

2. Measurements can be extended to the ultra-violet part of the spectrum using quartz lenses instead of glass ones and a suitable light source. That is ultra-violet rays can be used for excitation of fluorescence and resonance radiation as well in the ultra-violet part of the spectrum (intensity measured photographically or with a photoelectric cell or by means of excitation of secondary fluorescence) as in the visible one.

3. No polarizing prisms are necessary. The excitation can be therefore effectuated either by unpolarized or by polarized light. Using unpolarized light excitation we can measure the «true» duration of luminescence even in viscous media where the measured τ depends upon the directions of the exciting and returning electric vectors (7).

Finally it must be emphasized that analysing the «diffused light» of «mirror» curve we can deduce the laws of the transmission of light by such a quartz «valve» and check the theories of the diffraction of light by ultra-sound waves, for instance that of Raman.

We want to express our deep gratitude to S. I. Wawilow for his constant interest to our work. We are also indebted to S. M. Rytow for his kind cooperation.

Received
20. IV. 1937.

REFERENCES

[1] E. Gaviola, Ann. d. Phys., **81**, 681 (1926); ZS. f. Phys., **42**, 853 (1927); E. Duschinsky, ZS. f. Phys., **78**, 586 (1932); W. Szimanowsky, ZS. f. Phys., **95**, 440 (1935). [2] F. Duschinsky, loc. cit. [3] L. Brüninghaus, C. R., **204**, 118 (1937). [4] H. E. R. Becker, W. Hanle, O. Maerks, Phys. ZS., **11** (951), 414 (1936). [5] O. Maerks, Phys. ZS., **15** (955), 562 (1936). [6] W. Szimanowsky,

ZS. f. Phys., **95**, 450 (1936). [7] A. Jablonsky u. W. Szimanowsky, Nature, **135**, 582 (1935); W. Szimanowsky, ZS. f. Phys., **95**, 466 (1935); A. Jablonsky, ZS. f. Phys., **95**, 53 (1935); W. Kessel, ZS. f. Phys., **103**, 125 (1936).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

PHYSIQUE

VÉRIFICATION EXPÉRIMENTALE DE LA FORMULE DE THOMSON

Par K. S. LJALIKOV

(Présenté par S. I. Wawilow, de l'Académie, le 7. IV. 1937)

L'étude d'un nombre de faits physico-chimiques liés aux particules de la matière demande qu'on prenne en considération le changement de leurs propriétés, lequel dépend de leur dimension.

W. Thomson (1) a été le premier à indiquer l'influence de la courbure de la surface sur la force élastique de la vapeur, et plus tard Richard Helmholz et D. Gibbs y ont donné une base thermodynamique.

Vu la haute portée de l'équation de Thomson dans les diverses régions des sciences naturelles sa vérification expérimentale est d'un grand intérêt. Elle est bien difficile; d' après Thomson lui-même, la réalisation de l'expérience spéculative dont il se servait en déduisant l'équation serait à peine possible, puisque la distillation à faire du liquide qui passe d'un tube capillaire dans un autre n'aura lieu que bien lentement ce qui dépend du niveau du liquide.

En effet, les expériences effectuées par Bigelow et Trimble (2) ont démontré que le mouvement du niveau dans les tubes capillaires se produit dans la direction voulue, mais avec une telle lenteur que ce serait à peine possible d'établir l'état de l'équilibre.

Les essais de Shereshevsky (3) afin de déterminer l'élasticité de la vapeur d'eau dans des tubes capillaires ont complétement échoué. Ainsi, quoique la formule fût trouvée, il y a plus de soixante ans, elle n'a pas été une seule fois vérifiée.

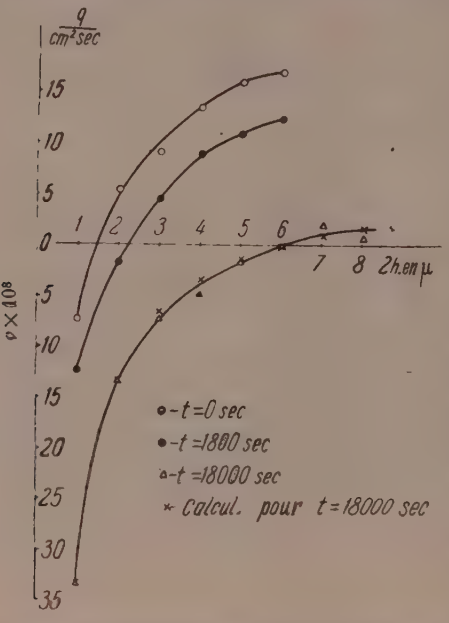
Vu que dans mes travaux sur la récristallisation de AgBr j'ai pris pour base la formule de Thomson, j'ai essayé de la vérifier dans un cas net autant que possible en employant la méthode élaborée pour l'examen du procédé de récristallisation.

La méthode consiste à mesurer la vitesse de l'évaporation de petites gouttes du liquide par rapport à leur rayon. La tentative d'appliquer cette méthode a été faite par Bigelow et Trimble eux-mêmes, cependant ils n'ont obtenu que des résultats peu précis, dont ils n'ont pu tirer aucune conclusion.

Les conditions de l'expérience sont les suivantes: on fait distiller une goutte de mercure, préalablement purifiée, dans une boule aux minces parois, bien rincée et dégasée dans le vide, à la temperature du début d'amolissement du verre. Puis la boule est soudée. Avant de procéder directement

à l'expérience on refroidit une partie de la paroi de la boule par l'air liquide, ce qui fait précipiter là une mince couche du mercure, laquelle après que l'air liquide soit éloigné se detache en nombre de gouttes de mercure dont quelques unes sont de la dimension d'une partie du micron et d'autres de plusieurs microns. Ensuite, on met la boule dans un thermostat rempli d'eau à la température de chambre et pendant plusieurs heures un nombre de microphotographies est effectué. Une fois agrandies elles permettent de compter la quantité de gouttes de chaque dimension quant au différents moments.

On peut calculer la vitesse de l'évaporation des gouttes de mercure d'après les courbes de la distribution des gouttes conformément à leur dimensions en se servant de la méthode appliquée par Tcherdyncev (4) dans nos recherces sur la croissance des cristaux de AgBr. Comme la vitesse de l'évaporation dans le vide est liée directement à la force élastique de vapeur ces données peuvent être employées pour vérifier la formule de Thomson.



Ci-après sont donnés les résultats du calcul des gouttes sur un espace de 0.074 mm². Les mesures ont été faites sur des photographies 500 fois agrandies. Toutes les gouttes ont été divisées en classes d'après leur dimensions. Le numéro de la classe désigne le diamètre de la goutte en microns.

L'expérience a été contrôlée par la constance de la masse du total de toutes les gouttes, laquelle restait constante dans les limites de l'erreur ce qui montre qu'au cours de l'expérience l'égalité de la température de toutes les parties du réservoir a été suffisamment bien maintenue, et l'effet observé ne dépendait pas de la distillation d'un endroit à un autre. Cela a été clair aussi par le fait que, sur un petit espace de 0.074 mm², les grandes gouttes augmentaient tandis que les petites diminuaient.

D'après les données sus-mentionnées les vitesses de l'agrandissement des gouttes ont été calculées en g/cm² sec. (table 2 et figure).

Toutes les valeurs φ sont multipliées par 10^8.

On peut admettre que la vitesse φ du croissement de la goutte du rayon r pour le moment donné est proportionnelle à la différence entre la force élastique de la vapeur $\bar{p}$ dans le réservoir et celle de la vapeur p_r au-dessus de la surface de la goutte:

$$\varphi = k(\bar{p} - p_r). \tag{1}$$

Introduisons dans l'équation (1) p_∞, la force élastique de la vapeur au-dessus de la surface plane:

$$\varphi = k(p_\infty - p_r) + k(p - p_\infty). \tag{2}$$

Table 1

Temps compté dès le début de l'expérience (en secondes)	Classes											Le total des gouttes
	1	2	3	4	5	6	7	8	9	10	11	
0	31	137	209	249	206	127	36	11	2	—	—	1 018
1800	38	113	184	221	215	154	63	18	2	—	—	1 009
18000	24	54	93	135	145	145	91	37	10	1	1	736

Table 2

Temps compté dès le début de l'expérience (en secondes)	Classes							
	1	2	3	4	5	6	7	8
0	− 7.31	+ 5.39	+9.09	+13.6	+16.1	+17.1	—	—
1800	−12.3	+ 1.75	+4.41	+ 8.95	+10.9	+12.5	—	—
18000	−33.3	−13.5	−7.18	− 4.77	− 1.53	0	+2.02	+0.806
18000	(−33.3)	−13.34	−6.66	− 3.32	− 1.33	(0)	+0.97	+1.67

D'autre part, selon la formule de Thomson nous avons:

$$p_r = p_\infty e^{\frac{2\sigma M}{RTd_{\text{ж}}r}}.$$

Décomposons $e^{\frac{2\sigma M}{RTd_{\text{ж}}r}}$ en une série et nous contentant de ses premiers membres seulement, $\frac{2\sigma M}{RTd_{\text{ж}}r}$ étant très petite, nous obtenons:

$$p_r - p = \frac{2\sigma Mp}{RTd_{\text{ж}}r}, \qquad (3)$$

$$\varphi = -k\frac{2\sigma Mp}{RTd_{\text{ж}}r} + k(\bar{p} - p). \qquad (4)$$

Dans l'équation (4) nous connaissons toutes les grandeurs excepté k et $\bar{p}$. En les calculant d'après les données expérimentales pour t=18000 sec., pour r=0.5 μ et r=3 μ, nous obtenons k=2.98·10^{-5} et $\bar{p}$=1.2422.

Ayant calculé d'après l'équation obtenue par ce procédé la valeur pour les autres points de la courbe de vitesse de l'accroissement des gouttes pour t=18000 sec., nous obtenons une coïncidence bien satisfaisante aux données expérimentales ce qui prouve qu'il y a une coïncidence de la force élastique de la vapeur au-dessus des gouttes de différentes dimensions à celle exigée par l'équation de Thomson.

Il est à noter que la valeur k est assez rapprochée de la valeur exigée par la théorie cinétique des gaz. Les conditions choisies par moi au cours de l'expérience ont les avantages suivants en comparaison des expériences de Bigelow et Trimble et de Shereshevsky: 1) il est facile d'étudier la conduite des globules aux dimensions minimes: on n'est limité que par la vitesse de leur évaporation et par la puissance du microscope; 2) grâce à la méthode statique les fluctuations possibles sont éliminées.

Institut optique.
Leningrad.

Manuscrit reçu
le 7. IV. 1937.

LITTÉRATURE CITÉE

[1] W. Thomson, Phil. Mag. [4], **42**, 448 (1871). [2] S. L. Bigelow, H. M. Trimble, J. Phys. Chem., **31**$_2$, 1798 (1927). [3] J. L. Shereshevsky, J. Am. Chem. Soc., **50**, 2966 (1928). [4] C. S. Lialikov, N. N. Piscounova, J. P. Chipilov, S. V. Tcherdyncev, IX Congrès Intern. Phot. Scient. et Appl., Paris. 1935.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

CHEMIE

KAINIT UND POLYHALIT IN DEN SALZLAGERN DER SOWJETUNION

Von N. S. KURNAKOV, Mitglied der Akademie, G. B. BOKIJ und I. N. LEPEŠKOV

Das Institut für allgemeine und anorganische Chemie der Akademie der Wissenschaften der UdSSR führt seit einer Reihe von Jahren unter der Leitung des Akad. N. S. Kurnakov wissenschaftliche Forschungsarbeiten über die Kalivorkommen der Kaspischen Tieflande.

Von den physikalisch-chemischen Expeditionen, die von dem Institut in den Jahren 1933—1935—1936 ausgeführt wurden, wurde ein reichhaltiges Material zur Charakteristik der Kalilager von West-Kasakhstan und des Saratower Gebiets gesammelt.

Von grösstem wissenschaftlichen und industriellen Interesse sind die Kaliablagerungen im Osinski Rayon ([1], [2]) (Saratower Gebiet) und im Temirski Rayon ([3]) (Aktjubinski Gau). Im Gebiet des Inder-Sees (West-Kasakhstan) wurden ausser den Kaliablagerungen auch Borate entdeckt. Kalisalzablagerungen gibt es auch im Sagisgebiet (West-Kasakhstan) 120 km südöstlich von Gurjew.

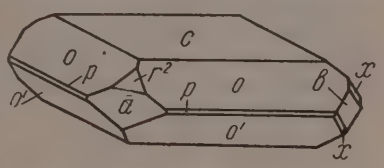


Abb. 1. Kainit aus Stassfurt nach Groth.

Bei den chemischen und kristalloptischen Untersuchungen der Salzproben, die in verschiedenen Tiefen der obengenannten Salzlager entnommen wurden, wurden neue Mineralien entdeckt, welche noch nie in Salzablagerungen der Sowjetunion begegnet waren. So wurde von uns zum Beispiel in den Salzproben, die aus den Kaliablagerungen von Osinki aus einer Tiefe von 587 und 590 m genommen wurden, im Jahre 1935 Kainit entdeckt.

Kainit ist ein Hydrat des Doppelsalzes von Magnesiumsulfat und Kaliumchlorid ($KCl \cdot MgSO_4 3H_2O$). Es wird in den Salzablagerungen von Stassfurt gefunden, wo es gemeinsam mit anderen Kalisalzen die Kainitschicht bildet. Bekannt sind auch die Kainitlager in Galizien und in Wetersheln (bei Magdeburg).

Die Farbe des Minerals ist grau, gelb, rot; die Festigkeit 2. Das spezifische Gewicht beträgt 2.1. Zuerst wurde der Kainit aus den Stassfurter Lagern von C. Zincken ([9]) beschrieben. Er nannte das neue Mineral Kainit (καίνοδ: griechisch—neu). Fast gleichzeitig wurde dieses Mineral auch

von Huyssen ([10]) beschrieben. Im Jahre 1865 unterzog Philipp den Kainit einer vollständigen Analyse. Im 1868 analysierte Frank die Kainitkristalle, welche P. Groth ([11]) im 1869 kristallographisch untersuchte.

Die kristallographischen Konstanten, welche von Groth angegeben wurden, findet man bis auf den heutigen Tag in den meisten Lehrbüchern der Mineralogie als Charakteristik des Kainits (Abb. 1). Groth fand, dass der Kainit zu der prismatischen Symmetriegruppe der monoklinen Sigonie angehört:

$$a : b : c = 1.2186 : 1 : 0.5863;\ \beta = 94°55'$$

An den Kainitkristallen wurden folgende Formen beobachtet:

$$a\{100\},\ b\{010\},\ c\{001\},\ p\{110\},\ s\{210\},\ l\{310\},\ \pi\{980\},$$
$$l_2\{510\},\ l_3\{340\},\ l_4\{230\},\ l_5\{130\},\ t\{101\},\ r\{201\},\ n\{410\},$$
$$d\{021\},\ o\{201\},\ o\{111\},\ w\{111\},\ \xi\{131\},\ e\{334\},\ x\{131\},$$
$$\theta\{221\},\ w\{311\},\ \varphi\{311\},\ \zeta\{223\} \qquad \{4 \cdot 18 \cdot 7\}?$$

Die optischen Konstanten des Kainits, die im Nachschlagebuch von E. S. Lorsen und Berman ([12]) angegeben sind, sind folgende:

$$N_g = 1.516; \qquad N_p = 1.494;$$
$$N_m = 1.505; \qquad 2_v = -85°.$$

Die von den verschiedenen Autoren angegebenen Brechungsindizes unterscheiden sich von den angeführten um 0.003—0.004.

Die Farbe des Kainits, welcher in den Salzlagern von Osinki entdeckt wurde, ist blass-rosa.

Die Salzablagerungen in Osinki sind von der Art der Salzkuppeln, von denen im Kaspischen Tiefland mehr als 300 erforscht wurden. Beim Bohren der Bohrlöcher № 1, 2 und 2 (a) wurde hier eine Schicht chemischen Ablagerungen des alten Permer Sees aufgedeckt welche Gipse, Anhydride, Steinsalz und Kali-Magnesiumsalze vom Typ des Karnallit, Sylvinit und Kainit enthielt. Wegen der ungenügenden Proben, die aus den verschiedenen Tiefen genommen wurden, gelang es nicht genauer die Schicht der Kainitablagerungen zu untersuchen. Aus den erhaltenen Proben konnte das Vorhandensein von Kainit in einer Tiefe von 587 und 590 m festgestellt werden.

Wir führen hier die Daten der chemischen Analyse der Salzprobe aus der Tiefe von 590 m an.

Tabelle 1

Salze	Menge des Salzes in Gewichtsprozenten
$CaSO_4$	2.46
$MgSO_4$	39.49
NaCl	10.26
KCl	21.06
$MgCl_2$	8.29
Trockener Rückstand	81.07

Nach der Untersuchung auf die Doppelsalze erhalten wir:

Kainit ($KCl \cdot MgSO_4 3H_2O$) = 48.87%.
Karnallit ($KCl \cdot MgCl_2 6H_2O$) = 24.07%.
NaCl = 10.26%.
$CaSO_4$ = 2.46%.

Optische Konstanten des gefundenen Kainits:

$N_g = 1.519$; $N_p = 1.492$;
$N_m = 1.504$; $2_v = -85°$.

Das Vorhandensein aller übrigen Minerale, die durch chemische Analysen festgestellt worden sind, wird durch kristalloptische Untersuchungen bestätigt.

Es sei bemerkt, dass die aus der Tiefe von 501—506 m genommenen Salzproben in der Hauptsache Natriumchlorid (91.02% NaCl) enthielten.

In einer Tiefe von 621—659 m gibt es Sylvinitablagerungen mit einem Kaliumchloridgehalt bis zu 22%.

Aller Wahrscheinlichkeit nach ist der Kainit in den Salzlagern von Ozinki von sekundärer Bildung.

Bei der Verdampfung der Kalisoolen (Inder-See) ([4]) scheidet sich nach dem Ausscheiden von Natriumchlorid Kaliumchlorid aus, dann Karnallit, Magnesiumsulfat ($KMnSO_4 \cdot 6H_2O$) und schliesslich Magnesiumchlorid.

Akad. N. S. Kurnakov und V. I. Nikolajev ([5]) erhielten Kainit aus den Soolen des Saks-Sees als sekundäres Produkt, das sich nach der Umkristallisierung des Karnallits und des Magnesiumsulfats gebildet hat.

Die Tatsache, dass in den Salzlagern von Osinki Kainitablagerungen gefunden worden sind, ist sowohl von wissenschaftlichem als auch von grossem industriellem Interesse, da man nach einiger Bearbeitung Kaliumsulfat erhalten kann welches bekanntlich für viel landwirtschaftliche Kulturen (Tabak, Wein, Kartoffel, Zuckerrübe u. a.) das beste Düngemittel ist.

In den Salzablagerungen des Aktjubinskigebiets (Salzkuppel Aschtsche-Bulak) und in der Inderski-Salzkuppel (West-Kasakhstan) wurde Polyhalit entdeckt.

Polyhalit ist ein wässeriges Kalium-, Magnesium- und Kalziumsulfat ($K_2MgCa_2(SO_4)_4 2H_2O$).

Es wird in den Salzlagern in Deutschland, in den Alpen und Galizien gefunden. In Stassfurt gibt es eine Polyhalitschicht, wo die Polyhalit- und Halitschlichten einander abwechseln. Die Farbe des Minerals ist rot, weiss, gelblich, gräulich. Festigkeit 3—3.5. Spezifisches Gewicht 2.72—2.78.

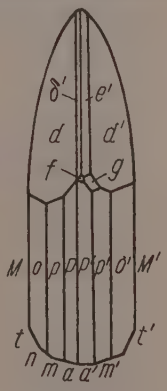


Abb. 2.
Polyhalit aus Stassfurt nach Görgey.

Zuerst wurde der Polyhalit aus den Lagern von Jschl von Stromeyer ([13]) beschrieben.

Die Polyhalitkristalle aus den Stassfurter Lagern wurden von R. Görgey ([14]) (Abb. 2) untersucht.

Durch Gonionmetermessungen erhielt er folgende Konstanten für den Polyhalit:

$$a : b : c = 0.9314 : 1 : 0.8562;\ \alpha = 92°29';\ \beta = 123°04';\ \gamma = 88°21';$$

er beobachtete folgende Formen:

$a\{100\}$, $M\{010\}$, $p\{001\}$
$\mu\{210\}$, $\lambda\{610\}$, $m\{2T0\}$, $l\{6T0\}$, $n\{230\}$, $\iota\{250\}$
$p\{011\}$, $o\{021\}$, $x\{101\}$, $h\{221\}$, $r\{211\}$, $\delta\{232\}$
$\varepsilon\{212\}$, $e\{212\}$, $d\{232\}$, $f\{214\}$, $g\{234\}$.

Gemäss der ersten Pinokoid gibt es eine Schichtung von mittlerer Qualität. Die Kristalle gehören der pinokoidalen Symmetriegruppe der triklinen Synogonie an.

Von ihm wurden auch die optischen Konstanten dieses Minerals gemessen ([15]).

Für Natriumlicht ist $N_p=1.548$; $N_m=1.562$; $N_g=1.567$; $2v=-70°$ (schwankt sehr).

Die von Görgey erhaltenen kristalloptischen und kristallographischen Konstanten wurden von den späteren Forschern bestätigt und werden heute als Charakteristiken des Polyhalits in allen Lehrbüchern der Mineralogie angeführt.

Bei der systematischen mikroskopischen Überprüfung der Salzproben aus den Salzlagern von Aschtsche-Bulak und bei der Messung der Brechungsindize der gefundenen Kristalle, richteten wir einerseits unsere Aufmerksamkeit auf die Homogenität der Stoffe, während man auf Grund der chemischen Analyse denken, dass wir es mit einem Gemisch von Kalium-, Magnesium- und Kalziumsulfat zu tun haben, wobei auch der geringe Gehalt an Kristallwasser charakteristisch ist. Anderseits wurde festgestellt, dass die gemessenen Hauptbrechungsindizes zu keinem der bis jetzt aus den Salzlagern der UdSSR bekannten Mineralien passten.

Später wurde gefunden, dass die gemessenen optischen Konstanten genau auf die Konstanten des Polyhalits * passen.

In den Salzlagern von Aschtsche-Bulak wurde der Polyhalit in zwei Zonen gefunden; in den Tiefen von 354—358 m und 500.5—536.3 m.

Wir führen die Daten (Tab. 2) der chemischen Analysen der Salzproben aus den oben genannten Tiefen des Bohrlochs № 1 in Aschtsche-Bulak an, wo der Polyhalitgehalt in den Salzproben bis zu 27% erreicht.

Tabelle 2

Tiefe in Metern	Salzmenge in Gewichtsprozenten						
	$K_2MgCa_2(SO_4)_4 \cdot 2H_2O$	$MgSO_4$	KCl	Na_2SO_4	NaCl	K_2SO_4	Unlösl. Rückstand in H_2O
354—358	21.80	12.97	15.64	13.06	2.26	—	31.25
525.1—536.3	27.37	21.91	14.19	—	2.98	0.71	30.53

Daten der kristalloptischen Untersuchungen des Polyhalits aus der Salzkuppel von Aschtsche-Bulak:

Tabelle 3

Tiefe in Metern	Optische Konstanten			
	N_g	N_m	N_p	$2v$
354—358	1.566	1.560	1.548	—66°
525.1—536.3	1.556	1.559	1.547	—70°

* In der Literatur gibt es zwei Hinweise auf das Vorhandensein des Polyhalits in der UdSSR: 1) im Solikamscher Gebiet ([6]) und 2) im Wolynskgebiet ([7]); aber späterhin wurden diese Hinweise nicht durch experimentelle Daten bestätigt, und es muss gesagt werden, dass wir zum erster Male in der UdSSR nur in der Permer Salzablagerungen des Ural-Emba Rayons dieses Mineral antreffen.

Die Farbe des Polyhalits ist rot.

In den Proben befindet er sich mit den Anhydriden, Kochsalz und Sylvin in Paragenese.

Bei der Bohrung des Bohrlochs № 1 in Aschtsche-Bulak wurde in Tiefen von 80—600 m Salzvorkommen gefunden.

In Tiefen von 80—213 m lagert Kochsalz mit einem Kaliumchloridgehalt bis zu 3%.

In Tiefen von 215—354 m gibt es Ablagerungen von Sylvinit, Karnallit (Karnallitgehalt bis zu 30%) und Kaliumchlorid (bis zu 25%). Noch tiefer beobachten wir, wie schon oben gesagt wurde, 2 Zonen des Polyhalits in Tiefen von 354—358 m und 500.5—536.6 m.

In den Zwischenraum zwischen diesen Zonen sind Ablagerungen von Steinzalz mit geringem Gehalt an Kaliumchlorid.

Polyhalit wurde auch in den Salzproben aus dem Bohrloch in der Inder-Salzkuppel (8) (die Bohrung wurde von ZNIGRI ausgeführt) gefunden. Bei der Untersuchung der Salzproben aus den Tiefen von 62 bis 126 m wurde in einigen Tiefen, z. B. von 96—99.2 m und 122—123 m, neben Steinsalz, Kaliumchlorid und Anhydriden bis zu 22% Polyhalit gefunden.

In Tab. 3 führen wir die chemischen Analysen der Salzproben des Bohrlochs № 17 der Inder-Salzkuppel aus den Tiefen von 96—99.2 m und 122—123 m an.

Tabelle 4

Tiefe in Metern	Salzmenge in Gewichtsprozenten									
	$CaSO_4$	$CaCl_2$	$MgSO_4$	$MgCl_2$	Na_2SO_4	NaCl	KCl	$K_2MgCa_2(SO_4)_4 2H_2O$	Unlösl. Rückstand	Trockene Rückstand
96—99.2	—	—	—	—	—	61.27	15.14	21.37	1.66	98.26
122—123	—	1.71	—	0.87	—	27.67	1.87	21.16	45.14	99.67

Daten der kristalloptischen Untersuchungen des Polyhalits aus den Tiefen von 96—99.2 m und 122—123 m.

Tabelle 5

Tiefe in Metern	Optische Konstanten		
	N_g	N_m	N_p
96—99.2	1.564	1.560	1.547
122—123	1.568	1.560	1.549

Der Winkel der Aussenachsen konnte wegen der Kleinheit der Kristalle nicht gemessen werden.

Der Polyhalit aus den angegebenen Lagern ist analog dem in Aschtsche-Bulak entdeckten Polyhalit roter Farbe. Wir werden uns nihct ausführ-

licher bei der Charakteristik der oben erwähnten Salzlager aufhalten, da hierüber ausführlich genug in der einschlägigen Literatur, die am Ende des Aufsatzes angeführt ist, beschrieben worden ist.

Es sei bemerkt, dass sich die Salzablagerungen des Kaspischen Tieflands und des Gemeinen Syrts, als Folge der Austrocknung des alten Permer Sees, gebildet haben.

Aller Wahrscheinlichkeit nach werden Salze vom Type des Polyhalits auch in anderen Salzablagerungen des Kaspischen Tieflandes entdeckt werden.

Der Fund von Polyhalit- und Kainitablagerungen in den Salzlagern der Sowjetunion ist nicht nur von wissenschaftlichem sondern auch von industriellem Interesse.

Institut f. allgemeine und anorganische Chemie.
Akademie der Wissenschaften der UdSSR.
Moskau.

Eingegangen d. 11. IV. 1937.

ZITIERTE LITERATUR

[1] Н. С. Курнаков, И. Н. Лепешков, Сорена, № 7 (1936). [2] Н. С. Курнаков, Н. И. Буялов, И. Н. Лепешков, Ж. калий, № 9 (1936). [3] Н. И. Буялов, И. Н. Лепешков, Д. И. Рябчиков, Ж. калий, № 1 (1937). [4] Н. С. Курнаков, И. Н. Лепешков, Сб. «Бор и калий в Западном Казахстане» (1935). [5] Н. С. Курнаков, В. И. Николаев, Сб. I Всеукраинской конферен. (1931). [6] Н. И. Сургунов. Рудный вестник, III, № 1—4, 12—22 (1918). [7] Ю. М. Абрамович, Тр. Волынск. геол. партии. 57, Житомир (1925). [8] Н. И. Буялов, И. Н. Лепешков, Д. И. Рябчиков, Экспед. сборник СОПС, АН СССР (1937). [9] C. Zincken, N. Jahrbuch. f. Miner., 310 (1865). [10] Huyssen, Berggeist, 10, № 15, 67, 68 (1865). [11] P. Groth, Pogg. Ann. d. Phys., **197** (1869). [12] E. S. Lorsen a. H. Berman, The Microscopic Determination of Nonopogul Minerals, sec. ed. (1934). [13] Stromeyer, Sweigs. J. f. Chem. u. Phys., **29**, 389. [14] R. Gorgey, Tscherm. Miner. u. Petrogr. Mitt., **33**, 48—102 (1914). [15] Ibidem, **29**, 209 (1910).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

CHEMIE

DIE KALILAGER AUF DEM RECHTEN UFER DER WOLGA UND IN KALMÜCKIEN

Von V. I. NIKOLAJEV, O. K. JANATJEVA und V. D. POLJAKOV

(Vorgelegt von N. S. Kurnakov, Mitglied der Akademie, d. 2. IV. 1937)

Gravimetrische Angaben ([1—2]) weisen offenbar auf die Existenz einer ganzen Reihe von Salzkuppeln auf dem rechten Ufer der Unteren Wolga, südlich von Stalingrad bis fast nach Astrachan, hin, die sich tief bis in die Kalmückische Steppe erstrecken.

Im Gebiete der Tschernyi Jar, bei der Siedlung Solenoje Zaimischtsche wurden von Wostokneft in den Jahren 1935—1936 zwei Bohrlöcher von 470 und 500 m Tiefe gebohrt. Wegen Mangel an Mitteln wurden leider die Bohrungen nicht fortgesetzt.

Auf unserer Expeditionsreise im Sommer 1936, die von dem Institut der allgemeinen und anorganischen Chemie der Akademie der Wissenschaften d. UdSSR organisiert worden war, erhielten wir einige Muster von Proben aus den Bohrlöchern № 1 und № 2.

Um den Salzgehalt der lehmigen Proben aus den verschiedenen Tiefen zu bestimmen, laugten wir die Salze aus den Proben mit heissem destilliertem Wasser aus, und bestimmten danach die Salzzusammensetzung der trockenen Rückstände.

Wie aus den Tab. 1 und 2 klar ersichtlich ist, ist der Kalium-, Brom- und Borsäure-Ionengehalt in dem Salzbestand der Proben nicht kaspischen Ursprungs.

Die Kaliumionenkonzentration in der Salzzusammensetzung erreicht 3.35%, die Bromionenkonzentration—0.34% und die Berechnung des Sylvinanteils im trockenen Rückstand (Mischung NaCl+KCl) ergibt sogar 0.57% und 0.67% Br'.

Ebenso erreicht der B_2O_3-Gehalt 0.28% der Salzzusammensetzung.

Der klare permische Charakter dieser Zahlen tritt deutlich beim Vergleich mit den Zahlen des Kalium-, Brom - und Borsäureionengehalts in der Salzzusammensetzung des Kaspischen Sees hervor.

Nach gut übereinstimmenden Analysen des Kaspischen Meerwassers von Schmidt und Lebedinzev enthält der trockene Salzrückstand dieses Meerwassers 0.56% $K^{\cdot}$ und 0.06% Br'. Für die Seen des Wolgadeltas sind diese Zahlen zuweilen noch kleiner, besonders für Kalium. Der Kalium-

Tabelle 1

Zusammensetzung der von den Proben aus dem Bohrloch № 1 mit Wasser ausgelaugten Salze

№	Bohrlöchertiefe	Einsalzen der Probe. Die Summe d. Gewichtsproz. der ausgelaugt. Salze	Gewichtsprozente des trockenen Rückstandes													Gewichtsprozent Bromions bzg. der Summe Natriums und Kaliums Chloride u. Bromide
			$Ca(HCO_3)_2$	$Mg(HCO_3)_2$	$NaHCO_3$	$MgSO_4$	$CaSO_3$	$CaCl_2$	KBr	KCl	NaCl	Na_2SO_4	B_2O_3	Salzen-Summe	K^+	Br^-	
1	256	1.222	3.10	1.02	42.21	—	—	—	0.25	3.18	29.25	20.82	0.17	100	1.75	0.17	0.52
2	312	0.424	8.40	5.17	9.70	—	—	—	0.33	4.71	69.25	2.18	0.26	100	2.58	0.22	0.29
3	473	0.450	15.60	6.50	—	—	15.65	6.08	0.36	5.64	50.00	—	0.22	100.05	3.08	0.24	0.43

Tabelle 2

Zusammensetzung der von den Proben aus dem Bohrloch № 2 mit Wasser ausgelaugten Salze

1	48.51	0.35	9.72	Spuren	—	Spuren	4.00	—	0.29	1.79	63.72	20.29	0.19	100	1.03	0.20	0.30
2	102—110	0.46	4.99	5.64	—	4.12	—	—	0.40	5.03	50.70	28.84	0.28	100	2.77	0.27	0.48
3	144.5	0.66	7.01	0.14	—	11.37	—	—	0.43	4.82	52.74	23.22	0.27	100	2.67	0.29	0.50
4	206—208	0.23	13.95	1.61	—	4.97	—	—	0.50	4.02	59.13	15.57	0.25	100	2.27	0.34	0.53
5	248—252	1.120	6.96	12.97	20.29	—	—	—	0.26	3.86	26.59	28.86	0.21	100	2.08	0.18	0.58
6	289—299	1.186	9.83	0.66	22.67	—	—	—	0.29	3.51	35.97	26.87	0.20	100	1.93	0.20	0.50
7	379—380	0.647	4.00	4.29	31.90	—	—	—	0.29	5.12	42.14	12.15	0.17	100.06	2.94	0.20	0.42
8	445	0.502	6.43	3.43	38.50	—	—	—	0.37	5.91	30.85	14.30	0.22	100.01	3.35	0.25	0.67

ionengehalt in den Soolen kann wegen der Adsorption durch den Seeschlamm bis zu einigen Hundertsteln eines Prozents betragen.

Vorläufig haben wir noch keine Analysen über den B_2O_3-Gehalt im Kaspischen Meerwasser. Aber aus vereinzelten Bestimmungen fanden wir, dass der B_2O_3-Gehalt in den Seesalzsoolen einige Hundertstel eines Prozents nicht übersteigt.

Die Zahlen: 3.35% $K^{\cdot}$, 0.34% Br' und 0.28% B_2O_3 wiederholen nur die Konzentration dieser Ionen in der Salzzusammensetzung des Inder-Sees ([3–4]).

Somit haben schon die physikalisch-chemischen Untersuchungen das Vorhandensein von Kali-, Brom- und Borsäuresalzen, wie auch von Salzen der Schwefel- und Kohlensäure (Tab. 1 und 2) in der Salzkuppel Solenoje Saimischtsche, die noch nicht durch Bohren geöffnet worden ist, festgestellt.

Allen Anzeichen nach haben wir es hier mit einer glücklichen Übertragung der Inder- und überhaupt der Ural-Emba-Bedingungen in das bevölkerte Wolgagebiet, mit einer Wasserader—der Wolga—in unmittelbarer Nähe und in geringer Entfernung vom Gebietszentrum Stalingrad, zu tun.

Die Fortsetzung der Bohrarbeiten in diesem Rayon zur Aufdeckung der Salzkuppel ist von dringender Notwendigkeit.

Hinweise auf das Vorhandensein von Kalilagern gibt es auch etwas südlicher von Solenoje Saimischtsche bei der Siedlung Kopanowka.

Eine Probe, die von dem Geologen J. S. Eventov aus einer naphtahaltigen Quelle bei Kopanowka entnommen und uns liebenswürdigerweise nach Moskau gebracht wurde, zeigte bei der Analyse und bei der Umrechnung auf den trockenen Rückstand: $K^{\cdot}$=2.14%, Br'=0.32%, B_2O_3=0.21%, d. h. die gleichen Zahlen wie bei der Salzzusammensetzung der Proben von Solenoje Saimischtsche.

Die physikalisch-chemischen Untersuchungen sprechen für das Vorhandensein von Kali-Tiefenstrukturen auch in der Kalmückischen Steppe nordwestlich von der kalmückischen Siedlung Jaschkul, längs der Sarpinischen Seenkette in eine Breite von 47° und nördlicher.

Die Beobachtungen des Astronomischen Instituts ([1]) ergaben hier negative Anomalien der Schwerkraft.

Wasserproben aus den Khuduken (Brunnen von 5—7 m Tiefe), die in dem angegebenen Gebiete zwischen 46 und 47° nördlicher Breite und in einer Breite von 47° liegen, haben bei den Analysen einen plötzlich erhöhten Gehalt an Kalium- und Bromionen gerade in der Breite von 47° bei allmählichen Übergang dieser Grösse zwischen 46 und 47° nördlicher Breite vom Kaspischen nach dem Permischen zu gezeigt.

Dies ist aus folgender Tab. 3 zu ersehen.

Tabelle 3

№	Bezeichnung des Brunnens	Gewichtsprozente des trockenen Rückstands		
		$K^{\cdot}$	Br'	Bemerkung
1	Oljùnch	0.30	0.16	
2	Gasalà	0.12	0.09	
3	Schar-Sarýn	0.41	0.10	
4	Schottà No 1	0.42	0.12	
5	Bjukjuntà	0.05	0.17	
6	Djadjà	0.06	0.19	
7	Gulì	0.51	0.10	
8	Kirjultschà	0.73	0.40	47° Breite, 100 km nördlich und etwas westlich von Jaschkul
9	Schottà No 2	2.90	0.24	

Die Weiterführung der Bohrarbeiten ist auch in diesem Gebiete in der nächsten Zeit von dringlicher Bedeutung.

Die zahlreichen Seen der Kalmückischen ASSR sind reich an schwefelsauren Magnesium- und Natriumsalzen. Die Kalisalze können in Verbindung mit schwefelsauren Salzen schwefelsaures Kalium geben, was zur Düngung und zur Umwandlung der riesigen, vorläufig noch unfruchtbaren und unbesiedelten Fläche der Kalmückischen Steppe in eine der blühenden Kornkammern unserer Union so notwendig ist.

Institut für allgemeine und anorganische Chemie.
Akademie der Wissenschaften d. UdSSR.
Moskau.

Eingegangen d. 2. IV. 1937.

ZITIERTE LITERATUR

[1] Ф. Ф. Голынец, Геологические исследования в Калмыцко-Сальских степях (1932). [2] В. А. Сельский, Соляные купола и их связь с нефтью (1936). [3] Н. С. Курнаков и И. Н. Лепешков, Сб. «Бор и калий» (1935). [4] В. И. Николаев и И. Н. Лепешков, Ж. Калий, № 9 (1936).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

CHEMISTRY

HYDROXYLAMINE-PYRIDINE COMPOUNDS OF BIVALENT PLATINUM

By I. I. ČERNIAJEV, Corresponding Member of the Academy, and V. I. GOREMYKIN

The earliest investigations of complex hydroxylamine compounds of platinum were made by Lossen (1), Alexander (2) and others.

In studying the reaction of hydroxylamine and potassium chlorplatinite, Lossen found that hydroxylamine could take up co-ordinate points in the internal sphere, to form a compound of the tetramine series having the composition: $|Pt4Hx|Cl_2$*. The reaction follows the equation:

$$K_2|PtCl_4| + 4Hx \rightarrow |Pt\,4Hx|\,Cl_2 + 2KCl.$$

The reaction seems to be a two-stage one, the first substance to be formed being cis $|Pt2HxCl_2|$, a non-electrolyte, which, incidentally, has not yet been isolated; that substance then reacts with the hydroxylamine to form Lossen's tetramine. Chemically speaking, the hydroxylamine in complex compounds generally behaves as a neutral amine, taking up only one co-ordinate point and being connected with the central atom of platinum through nitrogen (3). Alexander succeeded in synthesizing a large number of compounds. The greatest scientific interest is afforded by the base $|Pt4Hx|(OH)_2$ obtained by him and which had the co-ordinate number (4) 6 assigned to it similarly to the acetonitrile compounds of platinum (5); this base has been classed among those compounds that do not fit in with the generally adopted scheme of bivalent platinum complex compounds.

Attempts to synthesize complex hydroxylamine platinum compounds in combination with other amines were made by a number of authors, but the only ones to solve the problem successfully were L. A. Čugajev and I. I. Černiajev (6).

The experiments made by them have shown the possibility of synthesizing new complex platinum compounds that contain hydroxylamine, in different combinations, together with ammonia and partly with pyridine. On these complexes they have observed phenomena of isomerism, proved the equivalence of co-ordinate points in a complex molecule with the co-ordinate number 4, and have also developed a method for the transition from cis- to trans-compounds. The object of this work was to synthesize mixed

* Hx denotes hydroxylamine NH_2OH.

hydroxylamine pyridine complex compounds of bivalent platinum. We thought that a detailed study of those compounds might bring out new facts relating to the stability of complexes produced by various ratios of pyridine and hydroxylamine in the internal sphere of the compound, as these amines belong to quite different classes in point of structure, nature, and chemical complex-forming activity. A study of the reaction of mixed hydroxylamine pyridine compounds of platinum with hydrochloric acid might afford some indication as to the degree of stability of the chemical bond of pyridine and hydroxylamine with the central atom of platinum.

Another question that was of interest to us was the influence exerted by the composition of tetramine on the physical and chemical properties of the derivatives, i. e. of chlorplatinites. If we succeeded in bringing about the substitution of hydroxylamine for pyridine in the $|Pt4Py|Cl_2$ compound, we should arrive at the same co-ordinate type $|Pt4A|Cl_2$, but with the pyridine completely replaced by hydroxylamine. Mixed hydroxylamine-pyridine complex compounds may be represented by the following series:

$$\underset{\text{I}}{|Pt4Py|Cl_2};\quad \underset{\text{II}}{|PtHx3Py|Cl_2};\quad \underset{\text{IIIa}}{tr|Pt2Hx\,2Py|Cl_2};$$

$$\underset{\text{IIIb}}{cis|Pt2Hx2Py|Cl_2};\quad \underset{\text{IV}}{|Pt3HxPy|Cl_2};\quad \underset{\text{V}}{|Pt4Hx|Cl_2}.$$

As regards the compound represented in this series, compound I has been known since the classical investigations of Jörgensen (7). Compounds IIIa was obtained by L. A. Čugajev and I. I. Černiajev. The complex compound V has already been mentioned, it had first been obtained by Lossen. As to the compounds II, IIIb, IV and their chlorplatinites, they have first been synthesized and studied by the writers of this article.

The chlorites $|PtHx3Py|Cl_2$, cis-$|Pt2Hx2Py|Cl_2$ and $|Pt3HxPy|Cl_2$ are colourless crystalline substances readily soluble in water or diluted alcohol, and less soluble in pure alcohol. The chlorides may also be precipitated by the usual mixture of alcohol and ether or any other solvent, such as acetone, methyl-ethyl-ketone, etc. Treatment with hydrochloric acid or the heating of these compounds causes the splitting off of two amine molecules and the formation of the respective diamines, the course of the reaction being in full agreement with Jörgensen's rule. The solutions of the chlorides, acidulated with a few drops of hydrochloric acid, form with the potassium chlorplatinite, crystalline precipitates of the respective chlorplatinites:

$|PtHx3Py||PtCl_4|$, $|Pt2Hx2Py||PtCl_4|$ and $|Pt3HxPy||PtCl_4|$.

The analyses of the chlorides and chlorplatinites for platinum, chlorine and nitrogen, as well as the determinations of molecular electrical conductivity, are in agreement with formulae. The structure of these complex compounds may be inferred from the manner in which they are obtained, and is confirmed by the products of their splitting up with hydrochloric acid.

1. It has been found that the stability of hydroxylamine-pyridine compounds of platinum depends on their composition and is increased in the direction of Lossen chloride $|Pt4Hx|Cl_2$.

2. The composition of tetramines affects the properties of the chlorplatinites (colour, shape of crystals, solubility, stability in heating, etc.).

3. The action of hydrochloric acid on the tetramines of this series is in full agreement with Jörgensen's rule.

4. The authors have obtained the chlorides $|PtHx3Py|Cl_2$, cis-$|Pt2Hx2Py|Cl_2$, $|Pt3HxPy|Cl_2$ and the chlorplatinites $|PtHx3Py|_2|PtCl_4|$, $|Pt2Hx2Py|_2PtCl_4|$ and $|Pt3HxPy|_2|PtCl_4|$, which have never been obtained before.

Institute of General and Inorganic Chemistry.
Academy of Sciences of the USSR.
Moscow.

Received
25. III. 1937.

REFERENCES

[1] Lossen, Lieb. Ann., **160**, 242 (1871). [2] Alexander, Lieb. Ann., **246**, 239 (1888). [3] И. И. Черняев, Изв. Плат. ин-та, **7**, 52 (1929). [4] A. Werner, Neuere Aufsuche auf d. Gebiete d. anorg. Chem. (1923). [5] A. Tschugaeff, W. Lebedinski, C. R., **161**, 563 (1915). [6] А. Чугаев и И. И. Черняев, Изв. Плат. ин-та, **1**, 29 (1920). [7] Jörgensen, Journ. f. prakt. Chem., **33**, 489 (1886).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

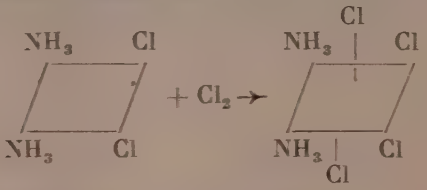
CHEMISTRY

OXIDATION OF HYDROXYLAMINE COMPOUNDS OF PLATINUM

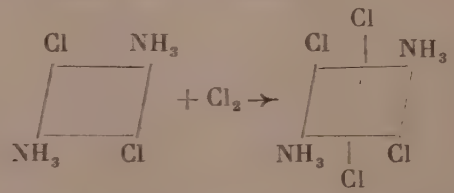
By I. I. ČERNIAJEV, Corresponding Member of the Academy, and V. I. GOREMYKIN

The problem of the oxidation of hydroxylamine compounds has not yet been systematically studied, probably because such a study could have hardly been expected to bring to light any exceptions to general rules. By analogy with other complex compounds of platinum, the oxidation process might have been expected to have the following results: the retention of all the other substitutes in the internal sphere, the setting up of a new co-ordinate, and the oxidation of hydroxylamine to nitrogen with its subsequent replacement.

A classical example of this is afforded by the following reactions: oxidation of Peyrone salt to Cleve salt



and oxidation of II Reiset salt to Gerhart salt:



As may be seen from these instances, not only are all the substitutes contained in bivalent platinum retained when the latter becomes oxidized but their relative geometrical positions remain unchanged. The same thing takes place in compounds of a much more complex composition if other oxidizers are used, e. g. hydrogen peroxide, bromine, diluted nitric acid, potassium permanganate, even these oxidizers fail to cause the amines to

split off or to change places. The object of the present work was to take advantage of the ready oxidability of hydroxylamine in order to substitute chlorine for it in tetravalent platinum. We were practically sure of getting, by means of the reaction:

$$2\,|Pt\,3NH_3Hx\,|\,Cl_2 + 3Cl_2 \rightarrow 2\,[Pt\,3NH_3Cl]\,Cl_3 + 2H_2O + 2HCl + N_2$$

from the rather easily obtainable tetramine $Pt3NH_3Hx\,|\,Cl_2$* the far less readily obtainable Cleve triamine $|Pt3NH_3Cl|Cl_3$.

It becomes, however, obvious from the very start that the process of oxidation of hydroxylamine compounds had some peculiarities of its own. To begin with, in the oxidation of hydroxylamine complexes, the platinum becomes oxidized after the hydroxylamine, which is abundantly clear from the fact the hydroxylamine can reduce the complex compounds of tetravalent platinum to bivalent platinum, and also from the impossibility to obtain a complex of tetravalent platinum containing hydroxylamine in the internal sphere. It is only this peculiarity that can account for another striking feature of the oxidation of hydroxylamine compounds, i. e. the destruction of the internal sphere in the process of oxidation, or rather the migration of some perfectly definite substitutes from the internal to the external sphere. To explain this migration of trans-groups we have to assume that the oxidizer, a molecule of chlorine or of hydrochlorous acid, first combines with the hydroxylamine to form one of two compounds: $NH_2OH{\cdot}Cl_2$, or $NH_2OH \cdot ClOH$. This unstable compound is converted into hydrochloric acid, gaseous nitrogen and water. Such excess of the oxidizer as remains thereafter reacts with the atom of platinum, causing the latter to become tetravalent. Still the complex molecule containing the hydroxylamine that has combined with the chlorine may exist for some time, and so long as it exists the hydroxylamine, it contains, has a strong trans-influence and can increase the mobility of the trans-substitute to such an extent that the latter may be replaced by water, hydroxil or chlorine, as the case may be.

A similar phenomenon had been previously observed by I. I. Černiajev and A. M. Rubinstein in the oxidation of Stromholm's $|Pt3NH_3SO_3|$ triamine, but, as only one compound had been investigated, we refrained at the time from any assumptions. At present, we have studied the oxidation of cis- and trans- $|Pt2NH_32Hx|Cl_2$ tetramines. If oxidized with chlorine in the cold, cis-tetramine gives, instead of diamine, a large yield of chlorplatinate of ammonium. The formation of the chlorplatinate of ammonium proves beyond doubt that the migration of ammonia molecules occurred at the moment of the oxidation of the hydroxylamine, since neither bivalent nor tetravalent cis-diamine can yield a chlorplatinate. Trans-tetramine gave tetravalent diamine, a normal product of oxidation, since no labilization of ammonia molecules could have occurred in the hydroxylamine during the process of its oxidation. Trans-diamine $|PtNH_3HxCl_2|$ also yielded chlorplatinate of ammonium on oxidation. Tetramine trans-$|PtHxNH_3{\cdot}2Py|Cl_2$ gave upon oxidation dipyridine-tetrachlorplatinate trans-$|Pt2PyCl_4|$.

Consequently, it is in the trans-position to hydroxylamine that the migration of ammonia takes place, irrespective of the number of hydroxylamine molecules in the internal sphere, or of the type of complex.

Triamine $|PtHxNH_3PyCl|Cl$ (trans-pyridine-chlorine) yielded, together with the normal product of oxidation (pyridine-amyne-tetrachloride of platinum), also a salt of Koss type $NH_4\,|PyCl_5Pt|$.

---

* Hx denotes hydroxylamine NH_2OH.

Finally, triamine $PtHxNO_2PyNH_3Cl$ (trans-hydroxylamine nitro-group) was converted into $PtPyNH_3Cl_4$ upon oxidation, i. e. the nitro-group was removed from the internal sphere.

The attempts to oxidize trans-diamine $[PtPyNH_3NO_2Cl]$ with chlorine have shown that in this compound the nitro-group does not oxidize, diamine $[PtPyNH_3NO_2Cl_3]$ being obtained.

Thus, the migration of the trans-substitute takes place even if the trans-substitute carries a charge; consequently, the process of migration does not depend on the nature of the substitute. It is rather curious that, contrary to the generally adopted view, the oxidation of a molecule may take place in a place other then that where the destruction of the molecule due to oxidation occurs. These two processes may be simultaneous, but spatially divided. We presume that the migration of trans-substitutes may be observed whenever the internal sphere substitute becomes oxidized before the central atom of platinum, i. e. in compounds of platinum with molecules similar to hydroxylamine, hydrazine, sulphur dioxide, oxalic acid, etc.

The investigation is being continued.

Institute of General and Inorganic Chemistry.
Academy of Sciences of the USSR.
Moscow.

Received
25. III. 1937.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

CHEMISTRY

ON THE REACTIONS OF RARE EARTHS AND ALLIED ELEMENTS WITH PYROGALLOL, GALLIC ACID AND MORPHINE. V.

By F. M. ŠEMIAKIN

(Communicated by N. S. Kurnakov, Member of the Academy, 25. III. 1937)

1. Reactions with Pyrogallol and with Gallic Acid

In addition to the reactions described in communication I (1), the reactions of praseodymium, erbium, yttrium and didymium (a mixture of neodymium and praseodymium) have been studied. The results obtained are tabulated below (Table 1).

Table 1

1% Gallic Acid in an Ammoniacal Medium

Compound	Facts observed
Praseodymium nitrate (Kahlbaum)	Bright pink precipitate. Grows lighter on standing
Didymium nitrate (Kahlbaum) (Pr+ +Nd)	Bright pink precipitate. Grows lighter on standing
Erbium nitrate (Kahlbaum)	Rose-coloured precipitate with a tinge of mauve. Grows lighter on standing
Yttrium nitrate (Poulain)	Yellow precipitate tinged with rose. Grows lighter on standing
Cerium (4) Sulphate prepared after Hillebrandt	Bluish-violet precipitate. Grows darker on standing
Thorium nitrate	Orange red solution, and flocculent precipitate of the same colour
Lanthanum nitrate	Brown solution, and flocculent precipitate of the same colour

The order of mixing was as follows:

Salt under examination, then gallic acid and finally concentrated ammonia solution.

In an ammoniacal medium, didymium nitrate gives with pyrogallol a yellowish-white precipitate. On standing, the precipitate becomes buff-coloured. The precipitates of neodymium and praseodymium have been described by Fernandes (2).

In an acetic acid medium, yttrium nitrate gives with gallic acid a pale yellow solution. In an acid medium, cerium sulphate (prepared after Hillebrandt) gives with gallic acid an evanescent pale organe colouring. The reactions were studied at 0.01—0.1 *M* salt concentrations. Thus, every one of these rare earth elements gives a characteristic bright precipitate or colouring. These reactions are of special interest in connection with the analyses of ores, such as lovchorite, loparite and others, which contain rare earth elements both of a basic and an acid nature (cerium, niobium, tantalum, etc.).

M. S. Platonov, N. F. Krivošlykov and A. A. Marakajev (3) have recently described new qualitative reactions on niobium and tantalum. It may be mentioned that these reactions, in so far as they concern polyphenols, have long been used in the separation of niobium and tantalum (4). As far as its reactions with pyrogallol and gallic acid are concerned, niobium shows a great similarity to titanium. An analogous reaction with titanium was used by the writers in a colorimetric determination (5).

In either a neutral or acetic acid medium, potassium tantalum fluoride gives no colouring with pyrogallol. In a hydrochloric or sulphuric acid medium, 0.1 *M* tantalum gives a rather faint sulphur-yellow colouring, a scanty precipitate being deposited on standing.

The tantalum salt used in these experiments was prepared by the writers from metallic tantalum dissolved in hydrofluoric acid, in the presence of a few drops of nitric acid or hydrogen peroxide, with a subsequent addition of caustic potash and sulphuric acid (after Treadwell-Hall) (6). Since pyrogallol becomes readily oxidized, it is advisable, in such experiments, to use gallic acid as a substitute (Table 2).

Table 2

1% Solution of Gallic Acid

Salt	Facts observed
K_2NbF_7	In a neutral medium: bright-yellow colouring. In an ammoniacal medium: orange-red colouring
K_2TaF_7	In a neutral medium: pale yellow. In an acetic acid medium: yellow. A precipitate is deposited
$ZrO(NO_3)_2$	In a neutral medium: colourless solution. In an ammoniacal medium: yellow-orange solution
Ti+Ce (4)	Titanium inhibits the reaction on cerium
Nb+Ce (4)	Niobium inhibits the reaction on cerium. Cerium neutralizes the reaction on niobium
Ta+Ce (4)	Tantalum inhibits the reaction on cerium. In ammonia, the solution is buff-coloured

Thus titanium, niobium and tantalum inhibit the reaction on cerium with pyrogallol and gallic acid. This explains why gallic acid fails to detect cerium in lovchorite ore, but may detect it in its concentrate (7).

2. Reactions with Morphine

In addition to communication III ([8]), the writers studied the reactions of morphine hydrochloride with salts of praseodymium, didymium (mixture of neodymium and praseodymium), erbium, yttrium, titanium, zirconium, niobium and tantalum. The results obtained are tabulated below (Table 3).

Table 3

Hydrochloric Morphine Solution

Salt	Facts observed
Praseodymium nitrate (Kahlbaum)	In ammonia gives no colouring at first; upon standing, however, the precipitate gradually becomes buff-coloured
Didymium nitrate (Kahlbaum) (Pr+Nd)	The same as praseodymium
Erbium nitrate (Kahlbaum)	In a neutral or ammoniacal medium gives no colouring with morphine, but only a flocculent white precipitate
Yttrium nitrate (Poulain)	The same as erbium
Titanium (4) sulphate (Kahlbaum)	The same as erbium
Zirconium nitrate (Kahlbaum)	White jelly-like precipitate
Potassium-niobium fluoride (Kahlbaum)	Fine white precipitate
Potassium-tantalum fluoride (from metallic tantalum)	Fine white precipitate

The dark brown colouring of cerium (4) sulphate with morphine in an ammoniacal medium gradually grows fainter, becoming buff-coloured. The colouring of praseodymium and didymium salts, on the other hand, intensifies with time, turning distinctly pale brown from hardly noticeable as it originally was. It is fainter, however, than the colouring of cerium (4) and cerium (3) with morphine and ammonia.

Niobium and tantalum do not prevent the reaction on cerium with morphine or ammonia, but weaken the resulting colouring. If an undiluted solution of potassium-tantalum fluoride prepared after Treadwell-Hall is made to react with morphine in an acid medium, a yellow colouring is obtained, which, in an ammoniacal medium, is converted into a fine, pale, rose-mauve precipitate. If the initial strong K_2TaF_7 solution is diluted with some water, no colouring is observed in either an acid or an ammoniacal medium. A fine, white, flocculent precipitate is deposited.

Conclusions

New reactions of rare earth elements with gallic acid, pyrogallol and morphine hydrochloride in an ammoniacal medium have been described.

Institute of General and Inorganic Chemistry.
Academy of Sciences of the USSR.
Moscow.

Received
25. III. 1937.

REFERENCES

[1] Ф. М. Шемякин, ЖОХ, **4**, 248 (1934); ZS. anorg. Chemie, **217**, 272 (1934). [2] Fernandes, Gazz. chim. Italiana, **56**, 682 (1926). [3] М. С. Платонов, Н. Ф. Кривошлыков и А. А. Маракаев, ЖОХ, **6**, 1815 (1936). [4] В. Гиллебранд и Г. Лендель, Практ. руков. по неорг. анализу, 519—520 (1936). [5] Ф. М. Шемякин и А. Неумолотова, ЖОХ, **5**, 493 (1935); Suell, Colorim. Analysis, 365 (1936). [6] Тредвелл-Голл, Курс аналит. химии, **1**, 595 (1933). [7] Ф. М. Шемякин и М. П. Владимирова, Заводск. лабор., **5**, 231 (1936). [8] F. Šemiakin, C. R. Acad. Sci. URSS, XIV, № 3, 113 (1937).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

GENETICS

INFLUENCE OF TEMPERATURE CONTRASTS ON THE FREQUENCY OF LETHAL MUTATIONS IN *DROSOPHILA MELANOGASTER**

By A. I. ZUJTIN

(Communicated by N. I. Vavilov, Member of the Academy, 9. IV. 1937)

If an organic form is adapted (all the other conditions being equal) to a certain average range of variations of some natural agent (temperature, humidity, pressure, etc.), the reaction of the organism as expressed by its life activity under such conditions may be termed an optimal one.

Any abrupt change in this external factor placing it beyond the average range of variations and bringing it nearer the maximum or the minimum will, if the new condition persists for some time, prove sublethal, and will result in the substitution of the pessimum of life activity for the optimum for the group of individuals concerned, possibly involving the ultimate destruction of most of them. A sharp unilateral change in such a factor, either towards the minimum, or towards the maximum, will produce a «semi-contrast» while a bilateral change involving a sharp rise up to the maximum followed by a sharp fall down to the minimum or vice versa, will produce a «full contrast» at different states of the factor.

Viewed from the standpoint of this scheme, the mutation effect obtained by various authors with high sublethal temperatures (Muller, 1928, 1932; Goldschmidt, 1929; Jollos, 1930—1933; Plough and Ives, 1932—1935; Timofeef-Ressovsky, 1935; Rokitski, 1931; Efroimson, 1932 and others), and by Gottschevsky with low sublethal temperatures (Gottschevsky, 1934), is a result of semi-contrast changes of the temperature factor. However, in most of the papers so far published on the effect of temperature semi-contrasts, such positive results as have been obtained have either not been quantitatively estimated with sifficient accuracy or have been found indefinite after such an estimate has been made. Moreover, some other authors, who used the same temperature effects failed to obtain any positive result (Ferry, Shapiro, Sidoroff, 1930; the unpublished results of Pedpield and Schulzt, Demerec and Sturtevaut, according to Muller's communication in 1932; Efroimson, 1930).

This lack of definiteness in the results obtained with temperature semi-contrast effects has induced the writer to employ f u l l t e m p e r a t u r e c o n t r a s t s as a factor of mutational variability (mutability).

* A brief abstract of a report delivered on April 8, 1936, at a meeting of the joint Scientific Seminar of the Chair Animal Genetics and the Chair of Plant Genetics, Leningrad State University.

Materials and Methods

The experiments were performed on a pure line of apricot-eyed *Drosophila melanogaster*, many generations of which bad been raised in our laboratory by breeding in larvae at different stages of development, from 1 to 6—7-day-old, as well as some young pupae were subjected to experiment. Before and after exposure, both the larvae experimented upon and the flies that had developed therefrom were kept in a thermostat at 25° C. The control cultures were kept in the same conditions. The insects were exposed to temperature contrasts described below. Full contrasts: *A*) (25°)→ 36° for 12 hrs. →0° for 12 hrs. →(25°); *B*) (25°)→0° for 12 hrs. →36° for 12 hrs. →(25°). Semi-contrasts: *A*) (25°) →36° for 12 hrs. →(25°), or (25°)→ →36° for 24 hrs.→(25°); *B*) (25°)→0° for 12 hrs. →(25°), or (25°) →0° for 12 hrs. →(25°).

Such lethal or semi-lethal mutations that had taken place either in the experiment, or in the control, were brought out by *ClB* method, and, consequently, only such variations were taken into account as had occurred in the *X*-chromosome. Visible mutations, mainly in lines with the lethal as well as less usual modifications, were incidentally noted, but no particulars of these are given in this paper. Neither are given the data on non-disjunction, the latter being recorded independently of the lethals.

Results of the Experiments

In accordance with the aim we set ourselves, attention was mainly centred on the effect of full bilateral contrasts. Table 1 summarizes the results of both (*A* and *B*) series of full contrast experiments. The age-groups have been arranged in series of 2-day intervals.

Table 1

Results of Experiments on the Influence of Full Temperature Contrasts on the Occurrence of Lethal Mutations in the *X*-Chromosome of Male Larvae of *Drosophila melanogaster* in the Apricot Line

Age of larvae	1—3 days	3—5 days	5—7 days	Total 1—7 days	Control
Number of *X*-chromosomes under investigation (F_2 culture)	2 199	1 995	902	5 096	3 488
Number of lethals and semi-lethals	5	9	6	20	3
Percentage of lethal mutations	0.227	0.451	0.665	0.392	0.086
Comparison of experiment with control	$\frac{0.141}{0.069} = 2.0$	$\frac{0.365}{0.088} = 4.1$	$\frac{0.579}{0.115} = 5.0$	$\frac{0.306}{0.077} = 4.0$	—
Comparison with control of combined middle and old age (3—7 days)	$\frac{0.518 - 0.086}{0.089} = \frac{0.432}{0.089} = 4.85$				

The comparison of experiment with control was performed by dividing the corresponding difference by its probable error calculated from the formula:

$$E_{m_{diff}} = 0.6745 \sqrt{\frac{a+b}{n_1 \cdot n_2}} \cdot 100\%.$$

From Table 1 we see first of all that different ages behave differently with regard to the factor under investigation. The old larvae are three times as mutable, and the middle aged ones are twice as mutable, as the young ones. A comparison of absolute percentages thus gives a ratio of 1 : 2 : 3 for the mutability of the three ages under consideration. A comparison of the relative values shown in the lowest line of Table 1 gives practically the same result.

The most essential result obtained is the fact that the mutation effect due to the action of full contrast has an unquestionable statistical significance. The old age of larvae gives a nearly 8fold increase of mutability as compared with the control, the significance and definiteness of the difference being adequately proved by the fact that it is 5 times as great as its possible error. The middle age of larvae shows a lower mutability than their old age; nevertheless it exhibits a 5fold increase in the frequency of mutation as compared with the control, the resulting difference being 4 times as great as its probable error.

The lowest line shows a comparison of the frequency of mutations observable in the middle age and in the old age combined, with the control data.

This comparison emphasizes once more the statistical significance of the results obtained.

Since, during the larval life of *Drosophila* mutability increases more or less steadily with age, the data obtained for its young age may be taken for the origin of an ascending regressive line. This shows that the young age fits in readily with the general scheme of mutability variations with age.

Table 2 shows a comparison of full contrast and semi-contrast ef-

Table 2

Results of the Comparison of the Mutation Effect in 5—6 Days Larvae, Due to Full Contrast *A* and Semi-Contrasts *A* and *B* of Different Duration (24 Hours and 12 Hours respectively)

Nature of exposure	Full contrast *A* (36°12 hrs, 0°12 hrs)	Semi-contrast *A* (36°24 hrs)	Semi-contrast *B* (0°24 hrs)	Semi-contrast *A* (36° 12 hrs)	Semi-contrast *B* (0°12 hrs)	Summarized results of all semi-contrasts	Control (25°)
Number of initial cultures (*P*)	20	55	50	10	10	125	25
Number of F_2-cultures investigated	59	433	468	159	315	1 375	1 113
Number of lethals and semi-lethals	1	1	1	0	1	3	1
Percentage of lethal mutations	1.695	0.231	0.214	0.000	0.317	0.218	0.090
Comparison of experiment with control	$\frac{1.605}{0.372}=4.3$	$\frac{0.141}{0.137}=1.0$	$\frac{0.124}{0.159}=0.8$	—	$\frac{0.227}{0.161}=1.4$	$\frac{0.128}{0.108}=1.2$	—
Comparison of full contrast effect with the combined effect of all semi-contrasts	$\frac{1.695-0.218}{0.473}=3.1$						

fects. Since the apricot line used in this set of experiments was of a different origin than those used in previous experiments, the control data relating to this line are shown separately from those given in Table 1. All the larvae used in this set of experiments were in their old age (5—6 days).

As may be seen from Table 2 a comparison of individual experimental data with the control data gives a statistically significant difference only with respect to the full contrast effect, as, on this case, the difference is 4,3 times as great as its probable error. As the number of recorded *X*-chromosomes was small and as only one mutation was produced by the full contrast effect, this result could not in itself have any serious importance, were it not for the fact that it is in full agreement with the data shown in Table 1. The similarity of the different semi-contrasts lies in the fact that none of them, either the 24-hours ones, or the 12-hours ones, gave any mutation effect essentially differing from the control data. The summarized results of all the four semi-contrasts, as shown in the penultimate vertical column, show a frequency of mutations twice as great as that appearing in the control column, but this difference lacks statistical significance. The full contrast effect, however, if compared with the summarized semi-contrast effects, shows, as may be seen from the lowest line of Table 2, as statistically significant difference in favour of a great efficacity of the full contrast.

A comparison of Tables 1 and 2 invalidates the suggestion that the full contrast effect might merely be the additive effect of two semi-contrasts.

Conclusion. The data cited above show that sharp contrast variations of temperature are a quite definite factor of mutability. This does not mean, of course, producing mutations. It is most probable that an abrupt change of any other factor in the surrounding medium (contrasts of humidity, pressure, etc.) might also produce hereditary modifications. Nevertheless, temperature variations stand first and foremost amongst all those contrast changes of external factors that may conceivably take place in nature.

There is no doubt that the validity of this proposition can only be proved if, side by side with the external factors, account is taken of those internal conditions of mutability that are inherent in the organism itself. That the latter may markedly effect the mutational efficacity of some external factor, is a well-known fact, and it is amply confirmed by the results of the present work on the dependence of mutation frequency on the age of the larvae. One would think that there was no necessity of introducing any unusual factors, such as radioactive energy, cosmic ray, artificial chemical substances, etc. to account for natural or spontaneous mutations, which ought to be adequately explained by those ordinary factors of the external medium that determine the development of the organic world provided the action of such factors is exercised under unusual or contrasting conditions.

Laboratory of Genetics and Experimental Zoology. Peterhof Biological Institute. Leningrad State University.

Received 9. IV. 1937.

REFERENCES

[1] W. Buchmann u. N. W. Timofeeff-Ressovsky, ZS. ind. A. V., **70** (1935). [2] L. Ferry, N. I. Shapiro, B. N. Sidoroff, Amer. Nat., **64** (1930). [3] R. Goldschmidt, Biolog. Zbl., **49** (1929). [4] G. Gottschewsky, ZS. ind. A. V., **67** (1934). [5] V. Jollos, Biol. Zbl., **50** (1930); Biol. Zbl., **51** (1931); Naturwiss, **19** (1931); Naturwiss., **21** (1933), u. a.: H. I. Muller, Genetics, **13** (1928); Proc. 6th Intern. Congr. Genet., I (1932). [6] H. H. Plough a. P. T. Ives, Genetics, **20** (1935). [7] Timofeeff-Ressovsky. ZS. ind. A. V., **70** (1935). [8] П. Ф. Рокицкий. Журн. экспер. биол., **6** (1930). [9] В. П. Эфроимсон, Журн. экспер. биол., **6** (1930); Биол. журн., **1** (1932).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

GENETICS

ON THE ELIMINATION OF SUPERNUMERARY CHROMOSOMES IN *ZEA MAYS*

By A. M. GROSSMAN

(Communicated by N. I. Vavilov, Member of the Academy, 13. IV. 1937)

The exact number of chromosomes in *Zea Mays* L. ($n=10$) was first determined by Kuwada ([5]), and was confirmed by subsequent investigations. However, for a number of varieties, particularly of sugary maize, pronounced deviations from this number have been found: by Kuwada ([5]) up to 25, by Randolph ([7]) up to 28, by Avdulov ([1]) up to 32, and, finally by Randolph (cited by McClintock) ([3]) up to 45.

The chromosomes in addition to the 10 normal pairs are called extra or supernumerary chromosomes ([6]), or chromosome types «B» and «C». They may be readily distinguished from the chromosomes of the basic set by their morphology and smaller size.

In his paper published in 1919 Kuwada reported that the number of chromosomes at meiosis in *Zea* varies in one and the same plant, in the case of plants whose chromosome number is more than 20 ($2n$). Kuwada explained this phenomenon, as also the presence in general of supernumerary chromosomes, as due to fragmentation of large chromosomes. Most of the later investigations accorded little or no attention to this remarkable phenomenon.

Consequently a solution to this theoretically very important problem as to the nature of the supernumerary chromosomes has not yet been even approximated by any investigator. Various theories have been advanced—fragmentation, duplication, irregular distribution of the chromosomes at anaphase, etc. But none of these theories give a real explanation of the nature of supernumerary chromosomes, because they do not take into account that these supernumerary chromosomes differ radically from the chromosomes of the basic set in their morphology, chemical composition, meiotic behaviour, and, most important of all, in their genetic constitution.

The supernumerary chromosomes may be distinguished from those of the basic set by their smaller size and by their terminal spindle attachment ([3]).

As regards differences in chemical composition between normal and supernumerary chromosomes, we may judge by their different reaction to staining.

The supernumerary chromosomes stain more deeply than normal chromosomes ([5,7]).

A characteristic peculiarity of supernumerary chromosomes is their heteropycnosis at prophases of meiosis ([3]).

At meiosis the supernumerary chromosomes remain univalents or conjugate with one another, sometimes forming multivalent groups ([3]). At the first division, when the chromosomes are separating to the poles, the supernumerary chromosomes usually lag behind the others, but occasionally they pre-

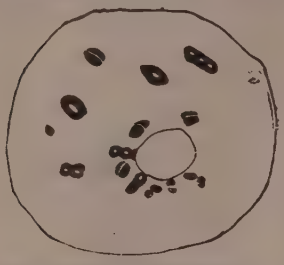


Fig. 1. Early diakinesis. 10 bivalents and 7 supernumerary chromosomes united in pairs.

Fig. 2. Diakinesis. 10 bivalents and 4 supernumerary chromosomes united in pairs.

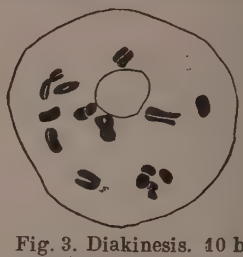


Fig. 3. Diakinesis. 10 bivalents and 5 supernumerary chromosomes; in addition, there is in the nucleus a very slightly stained body resembling a chromosome.

cede the rest. Sometimes they divide equationally at the first division, but most of them pass to the poles at random without splitting.

Randolph ([7]) already had sufficient grounds to conclude that the supernumerary chromosomes are genetically inert, since he found different numbers of such chromosomes in genetically identical material. However, Avdulov ([1]) was the first to formulate this conception precisely and concretely.

The data of Kuwada ([5]) regarding the inconstancy of the number of supernumerary chromosomes in the same plant give us further grounds for con-

Fig. 4. First metaphase. 10 bivalents and 2 supernumerary chromosomes.

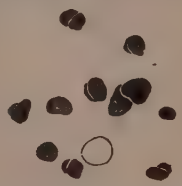


Fig. 5. First metaphase. 10 bivalents and 1 supernumerary chromosome.

Fig. 6. First metaphase (side view). 12 bivalents.

sidering the supernumerary chromosomes fundamentally different from the chromosomes of the basic set. However, Kuwada's ([5]) data have not been confirmed by any other investigator, and Kuwada ([5]) himself, by way of caution, states that the possibility of error in counting the number of chromosomes cannot be excluded.

In our present study of meiosis in *Zea Mays* (variety *Black Mexican*) we also encountered variation in the chromosome number in the same individual plant. By counting the number of chromosomes at various phases of the first division we discovered that the variation in the number of supernumerary chromosomes is due to the fact that some of them are eliminated bet-

ween early prophase and metaphase. Thus, for plant No. 33 we found at the stage of early diakinesis that the number of chromosomes was 27 (Fig. 1), while at middle and late diakinesis we counted 26, 25, and 24 chromosomes (Figs. 2 and 3). At metaphase there was observed a further diminution in the number of chromosomes—24, 23, and even 22 and 21 (Figs. 4, 5, and 6). It should be stated, however, that occasional metaphase plates were encountered with a large number of supernumerary chromosomes (Fig. 7), but in general the process of elimination of the supernumerary chromosomes from prophase to metaphase was quite definite. Metaphase plates were found with a small number of supernumerary chromosomes and several very slightly stained bodies resembling chromosomes in shape (Fig. 8).

Similar variations were found in a number of other plants. Thus, plant No. 15 had at early diakinesis 10 bivalents and 4 univalents, while at early

Fig. 7. First metaphase. 10 bivalents and 6 supernumerary chromosomes; 4 of the latter have formed a ring.

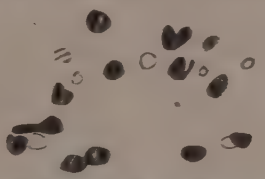


Fig. 8. First metaphase. 9 bivalents, 3 univalents, and 6 chromosome-like bodies very slightly stained.

metaphase there were 10 bivalents and only 3 univalents. In plant No. 3 at diakinesis 25 chromosomes were counted; at metaphase 24 and 23 chromosomes. In plant No. 31 at early diakinesis 10 bivalents and 3 univalents were counted; at late diakinesis 10 bivalents and 2 univalents; and so on.

In the light of our data on the elimination of the supernumerary chromosomes at meiosis the facts cited by Randolph (7) regarding the difference in number of chromosomes in the soma and at meiosis in the same plant become understandable. Thus, in one plant he found in the soma 26 chromosomes, and at meiosis 25; in another plant the respective numbers were 28 and 26. It may be assumed that Randolph (7) counted the chromosomes at meiosis after part of them had already been eliminated.

Longley (6) reported that he obtained from normal 20-chromosome parents five 22-chromosome plants and one 21-chromosome plant. If he determined the chromosome number of the parents at meiosis, then he may have made the chromosome counts at metaphases, when the supernumerary chromosomes had already been eliminated. However, gametes might have been formed including some of the supernumerary chromosomes. This would explain why Longley obtaines such progeny.

Somewhat unique are the facts cited by Fisk (2). In one adventitious root the chromosome number varied from 20 to 23, being 22 in the majority of cases. Might it not be that elimination sometimes occurs in somatic tissues as well?

Being unable to explain these phenomena, the investigators regarded them as «insignificant» (6) or evade this problem completely (1,4), although they cite Kuwada's (5) paper in their bibliographies.

The elimination of the supernumerary chromosomes during the period from the early to the late stages of the first division is additional confirmation of the assumption that the supernumerary chromosomes differ fundamentally in their nature from the chromosomes of the basic set.

This investigation was performed under the supervision of G. A. Lewitsky, to whom I desire to express my deep gratitude.

Cytological Laboratory.
Institute of Plant Industry.
Leningrad.

Received
13. IV. 1937.

REFERENCES

[1] Н. П. Авдулов, Тр. по прикл. бот., ген. и сел., серия 2, № 2 (1933). [2] E. L. Fisk, Proc. Nat. Acad. Sci., 11 (1925). [3] B. McClintock, ZS. f. Zellf. u. mikr. Anat., 19. No. 2 (1933). [4] Z. A. Kozhuchow, Planta, 19, No. 1 (1933). [5] Y. Kuwada, Jour. Coll. Sci., Imp. Univ. Токуо, 39, Art. 10 (1919). [6] A. E. Longley, Jour. Agr. Res., 35, No. 9 (1927). [7] L. F. Randolph, Corn. Univ. Agr. Exp. Sta., Memoir 117 (1928). [8] L. F. Randolph, The Anatom. Rec., 41 (1928).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

GENETIK

DIE CHROMOSOMEN IN DER SPERMATOGENESE DES HAUSESELS

Von I. SOKOLOV

(Vorgelegt von N. I. Vavilov, Mitglied der Akademie, d. 14. IV. 1937)

Über die Karyologie des Hausesels existieren bis jetzt keine Angaben. Von den übrigen Vertretern der Fam. *Equidae* wurde nur das Pferd untersucht. Aber auch n bezug auf dieses Objekt verfügen wir eigentlich nur über eine einzige Arbeit Painter's (1924), denn alle früheren Autoren (Kirillow, 1912, Wodsedalek, 1914, Masui, 1919) geben eine allzu niedrige Schätzung der Chromosomenzahl an.

Das vorliegende Material wurde von I. I. Novikov, Aspiranten der Leningradscher Universität, im April 1936 gesammelt und durch Kastration von zwei Eseln in der Pferdezuchtanstalt N. 68 Urta-Ssaraj, in Taschkent gewonnen. Beide Exemplare waren 4 Jahre alt und stammten von Müttern, die der weissen bucharischen Rasse angehörten; die Väter blieben unbekannt.

Kleine Testikelstückchen wurden nach Bouin-Allen und nach Champy fixiert. Wenn das erstgenannte Fixiermittel ausgezeichnete allgemeine Bilder des histologischen Baues der Testikel ergab, so erwies es sich im gegebenen Falle für zytologische Zwecke als nicht genügend. Viel bessere Resultate wurden mit dem gemisch von Champy erreicht. Die 7—8 dicken Schnitte wurden mit Eisenhämatoxylin nach Heidenhain gefärbt.

Spermatogoniale Metaphasen (auch Anaphasen) kamen bei beiden Exemplaren sehr oft zur Beobachtung. Leider ist es mir nicht gelungen, trotz Durchsicht einer grossen Anzahl derselben, völlig klare Chromosomenplatten zu erhalten. Wenn man sich auf Zählungen in den Spermatocyten stützt, so ist in den Spermatogonien als diploide Zahl 64 zu erwarten. In der Tat erhielt ich in den besten von mir beobachteten Platten, wo die Chromosomen nicht zu stark zusammengedrängt lagen, Zahlen, die sehr nahe bei 64 lagen (häufiger waren sie grösser als 64) (Abb. 5). Von der Morphologie der Chromosomen lässt sich im allgemeinen sagen, dass dieselben verschieden lange, meist gerade Stäbchen darstellen. Auf manchen Platten kommen auch einige wenige V-förmig gebogene Chromosomen zur Beobachtung, was jedoch kaum ihr charakteristisches Merkmal darstellt und wohl durch mechanische Ursachen hervorgerufen wird. In jedem Fall muss die Anwesenheit von echten V-förmigen, d. h. zweischenkeligen Chromosomen mit medialem Anhaften der Spindelfaser in Abrede gestellt werden. In dem Peripherieteil der Spermatogonialplatte liegen, wie gewöhnlich, vorwiegend die grössten Chromosomen, die kleineren

nahmen meist den zentralen Teil ein. Unter den peripliergelegenen Chromosomen fallen ins Auge 1—2 allergrösste Chromosomenpaare und ein solches unpaariges Chromosom, welches ich für *X*-Chromosom halten möchte. Das *Y*-Chromosom lässt sich durchaus nicht immer beobachten. In günstigen Fällen tritt es als das kleinste Chromatinelement auf, das keinen Partner hat.

Die Chromosomen der Metaphase der I. Reifungsteilung erscheinen bei Fixierung nach Bouin-Allen meist unregelmässig verteilt und zu mehrere Haufen oder Reihen zusammengeklebt, ganz so wie wir es bei Painter (l.c.) auf Abb. 11 12, 20 vorfinden. Die Fixierung nach Champy ergibt hingegen ziemlich regelrechte Chromosomenplatten, wo die Chromosomen nicht zusammenkleben und in annähernd gleichen Entfernungen von einander liegen (Abb. 1 u. 2). Gewisse Schwierigkeiten bei der Zählung der Bivalente rühren daher, dass einige von ihnen nicht selten übereinander lagern oder auch

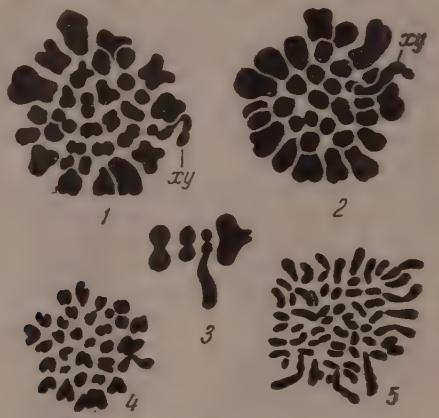


Abb. 1 u. 2. Zwei Metaphasen der I. Teilung.
Abb. 3. *X*-, *Y*-Paar nebst drei Bivalenten, von der Seite.
Abb. 4. Metaphase der II. Teilung.
Abb. 5. Spermatogoniale Metaphase.

ihre Orientierung in der Äquatorialplatte ändern, wobei bei hantelförmigen Bivalenten die beiden Partner entweder sich schief zur Äquatorialebene einstellen oder sich in diese Ebene nebeneinander legen.

Die Zählung der Bivalente bei einer grossen Anzahl (über 70) I. Metaphasen ergab, wie es in solchen Fällen meist zu sein pflegt, eine gewisse Schwankung der Zahl zwischen 29—33. Die Zahl 29 wurde nur dreimal beobachtet, 30—in etwa 10—12% aller Fälle; die Zahl 33 wurde nur durch gewisse Interpretation der entsprechenden Metaphasen gewonnen. Zieht man die obenerwähnten Abweichungen in der Lagerung der Bivalente als mögliche Fehlerquellen bei der Feststellung ihrer Zahl in Betracht, so lässt sich mit ziemlicher Bestimmtheit sagen, dass die haploide Chromosomenzahl beim Hausesel 32 betragen muss, zumal diese Zahl bei weitem am häufigsten beobachtet wurde (Abb. 1 u. 2).

Die einzelnen Bivalente differieren ziemlich stark in ihrer Grösse und ergeben eine Reihe mit allmählich abnehmenden Dimensionen. Der grösste Bivalent ist etwa 5—6mal länger als das kleinste. In der Äquatorialplatte gruppieren sie sich derartig, dass die grösseren Bivalente zu denen sich noch paar kleinere hinzugesellen, den Peripherieteil derselben zusammensetzen, während die kleineren den zentralen Teil einnehmen. Infolge der allmähli-

chen Grössenabstufung der Bivalente ist eine Identifizierung derselben auf verschiedenen Platten kaum möglich. Eine Ausnahme bilden die 2—3 grössten Bivalente, die an ihren grossen Dimensionen leicht kenntlich sind, und dann auch das Geschlechtschromosomenpaar.

Von der Anwesenheit eines ungleichen XY-Chromosomenpaares kann man sich leicht bei Seitenansicht der I. Metaphasen überzeugen, bei der die Bivalente ziemlich regelmässig in eine Ebene geordnet erscheinen. Dabei nimmt man ein langes Chromosom wahr, das sich nur mit seinem Proximalende in der Äquatorialebene befindet, mit dem anderen, dagegen, ausserhalb derselben liegt. Das ist das X-Chromosom (Abb. 3). Es kann entweder gerade oder verschiedenartig gekrümmt und mit seiner Längsachse verschieden gegen die Äquatorialplatte orientiert sein. In vielen Fällen beobachtet man an diesem Chromosom Einschnürungen, die dieselbe in etwa 5 ungefähr gleiche Abschnitte teilen. Am deutlichsten ist immer der proximalste Abschnitt abgegliedert, der in der Form eines kompakten Segmentes von ungefähr rundlicher Gestalt hervortritt und der sich in die Äquatorialplatte einstellt. Was das Y-Chromosom anbetrifft, dass das kleinste Element des Satzes darstellt, so befindet es sich immer in Verbindung mit dem X-Chromosom, und zwar mit dem proximalen Segment des letzteren, mit dem es seiner Grösse und Form nach homologisiert werden muss (Abb. 3). Obwohl Anaphasen der I. Reifungsteilung fast gar nicht beobachtet wurden, darf es immerhin aus der Lage des X- und des Y-Chromosoms in der Metaphasenplatte, wo sie gegen verschiedene Pole gerichtet sind, geschlossen werden, dass die I. Teilung für dieselben eine Reduktionsteilung ist. Somit muss es für den Hausesel als festgestellt gelten, dass sowohl der allgemeine Charakter der Geschlechtschromosomen, als auch ihr Verhalten während der Meiose völlig den Verhältnissen bei anderen Säugetieren, darunter auch beim Pferde, entsprechen.

Metaphasen der II. Reifungsteilung wurden ziemlich oft beobachtet, jedoch ganz einwandfreie Bilder kamen infolge der engen Gruppierung der Chromosomen sowie infolge der gegenseitigen Abstossung der Distalenden der Tochterhälften, was ihnen ein V-förmiges Aussehen gibt, nur vereinzelt vor. In denjenigen Fällen, wo eine Zahlung möglich war, erhielt man entweder die Zahl 32 (Abb. 4) oder eine ihr nahestehende Zahl.

In den Spermatozyten der I. sowie der II. Ordnung wird regelmässig der sog. «Chromotoide Körper» beobachtet, dessen Bedeutung und Funktion ganz rätselhaft bleiben. Er wird von Painter auch für die Spermatozyten des Pferdes erwähnt. Er hat eine kugelförmige Gestalt und lässt eine Differenzierung in die dunklere, ziemlich dicke Peripherieschicht und den blasseren Zentralteil erkenen.

Beim Vergleich unserer Befunde mit denjenigen von Painter beim Pferde treten gewisse Unterschiede zu Tage. So findet Painter in den Spermatogonien des Pferdes eine ziemlich grosse Anzahl V-förmig gekrümmter Chromosomen vor. Inwiefern, jedoch, sämtliche dieser Chromosomen wahre zweischenkelige und nicht gekrümmte einschenkelige Chromosomen sind, lässt sich ohne erneute Studien nicht entscheiden. Eine weit wichtigere Diskrepanz besteht in der Chromosomenzahl. Wenn man kein besonderes Gewicht auf die Zahlungen in der Spermatogonien legt, welche bei beiden Objekten nur annähernde Resultate (für den Esel etwa 64, für das Pferd etwa 60) ergaben, so erhalten wir doch für die Spermatozyten als Haploidzahlen: $n=32$ für den Hausesel und $n=30$ für das Pferd. Es wäre ja wohl sehr verlockend aus diesem Grunde die Sterilität der Maultiere auf Unterschiede in der Chromosomenzahl in den Chromosomensätzen der beiden Elternformen zurückführen zu wollen. Jedoch ist ein derartiger Schluss noch verfrüht, denn, meiner Mei-

nung nach, können die Ergebnisse Painter's infolge nicht ganz befriedigender Fixierung seines Materials (Bouin-Allen), die eine Verschiebung und eine teilweise Verklebung der Chromosomen herforgerufen hat (vgl. z. B. seine Abb. 1, 2, 11, 12 und 20), nicht als endgültig gelten. Aus diesem Grunde gedenke ich, parallel mit dem zytologischen Studien der Testikel der Maultiere, auch die Karyologie des Pferdes einer Nachuntersuchung zu unterziehen.

**Laboratorium für Genetik
und experimentelle Zoologie.
Staatsuniversität Leningrad.**

Eingegangen
d. 14. IV. 1937.

ZITIERTE LITERATUR

Th. S. Painter, Journ. exp. Zool., **39** (1924).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

PLANT PHYSIOLOGY

HORMONIZATION OF SEEDS—A POSSIBLE AGROTECHNICAL PROCESS

By V. I. TOVARNITSKIJ and T. L. RIVKIND

(Communicated by A. N. Bach, Member of the Academy, 31. III. 1937)

In the works of Shander (1934) and in a number of investigations conducted by N. G. Cholodny the rôle of the hormones of the seed in germination (1933—1935) was studied, and the means of their transport, as well as their influence on the vitability of the embryo were established. In his last two communications (1936),N . G. Cholodny put forward a scientific theory of yarovization and experimentally proved the possibility of introducing, in addition, hormones into seeds from outside, by soaking them in hormone-laden solutions. The experiments showed that the so-called hormonization of the grain, in addition to stimulating the early stages of its germination, exerts a certain after-effect of fairly long duration which shows itself in the acceleration of the general development of the superterranean mass, in the quickening of flowering and, as the ultimate result, in the increase of the yield. It was after the completion of our experiments that to our knowledge came the latest works of R. Pohl (1936) who made still further progress in this direction and proved the possibility of not only extracting hormones from the seed by various means, but also of their repeated artificial introduction. It is of importance to note that in addition to heterauxin raw extracts from urine were also used in these experiments, for hormonization of non-germinating seeds.

Basing our investigations on the same arguments as N. G. Holodny did, but bearing in mind that certain other hormonal factors may prove to be essential to plants, we carried out in 1936 a number of minor experiments on hormonization of the seeds of Spring wheat, vetch and sugar beet with urine and extract of yeast, in order to determine the practical possibility of actually increasing by this process the yield of the plants. We may mention that preliminary laboratory investigations enabled us to extend our experiments in 1936 directly to field conditions.

On the basis of a number of works, we used in our investigations in 1936 the urine of foal and non-pregnant mares, as a hormone-containing extract, taking as a starting point in our experiments the following evidence:

1. Kögl proved the presence in urine of a number of vegetable hormones: auxins a, b, and heterauxin (about 2mg to a liter), bios, substances of the risoline type and those of group B. The importance of all these substances to plants has been proved byy a considerable number of works.

2. The presence of animal hormones in urine, both of the sexual (phollikulin, testikulin, equilin, equilinin and hippulin) and the hypophysar series, has been proved. There is a wide literature on the effect of these hormones on plants, but as the evidence given is contradictory, no final conclusions may be drawn. The favourable effect of equilin and equilinin on the plant is observed in the works of M. Janot (1934).

3. M. Roscoe (1936) proved the presence in urine not only of vitamins A and C, but also those of Group B (B_1 and B_2), the favourable effect of which on the plant has been repeatedly observed.

4. Urine contains sterins, which are related to certain hormones and vitamins, as well as flavins (uroflavin, Koshara, 1935) the effect of which on the plant has not yet been thoroughly studied but which, it appears, are also capable of producing some «stimulating» influence and from which the vegetable organism itself can possibly synthesize the compounds it needs.

Thus, as is seen, urine represents a source of a wide variety of vitally necessary substances of the nature of hormones or that closely related to it, and the problem would, therefore, consist in finding the most practicable means of their utilization.

Taking as a basis the works of N. Nielsen (1931), B. Nath (1934), A. Virtanen (1933) and von Hausen on yeast, as well as a number of other works, by which, in connection with the presence in yeast of substances of Group B, bios and vitamine B_1 and B_2, their favourable effect on higher plants had been established, we also used in our studies on hormonization of seeds aqueous extracts (when boiling) of fresh bread yeast.

Experimental

The experiments of 1936 were conducted in field conditions at the experimental base of the Institute in the suburbs of Moscow, on field plots of different sizes: a) with wheat of field plots of 1 sq. meter, 4 times repeated, and those of 50 sq. meters, 3 times repeated; besides, a part of the experimental work was organized at the collective farms on field plots of 250—1 000 sq. meters, twice repeated; b) with vetch on plots of 1 sq. meter, 4 times repeated, and c) with beet on plots of 1 sq. meter, 20 roots on each plot, 4 times repeated.

Taking as a starting point the same ideas on which N. G. Cholodny based his studies, we treated the seeds with hormone-containing solutions, by two methods: in one case by simply soaking the seeds in the solutions, and in another with subsequent yarovization.

Of great and conclusive importance in the matter of hormonization of seeds with such hormone-containing solutions as we used in our experiments, is the concentration of the solutions. Only diluted solutions are to be applied as the concentrated ones render the seeds ungerminative and destroying their vitality.

Of the first favourable fact which manifested themselves in our experiments, are to be mentioned the aseptic properties of urine and the good protection of the seeds from fungi, under conditions of yarovization. Further, hormonization showed itself in the vigour of shoot growth (an increase of up to 30%) and in the general development of the plants of wheat.

No marked differences in the commencement of individual stages of development of the plants were observed. This may have been due to the drought conditions of 1936 which brought about a quicker maturing of the plants (by 1—1½ months), and a considerably reduced yield. In experiments on hormonization of the seeds of wheat with the urine of mares in foal the following results were obtained on micro-plots:

Table 1

Field plots	Experiment. field			Collect. farm «International»
	At harvest time			In the period of earing
	Weight of 100 plants in g	Length of ear in cm	Thickness of ear in cm	Weight of 100 plants in g
Control (sowing with dry seeds)	168	5.7	7.3	30
Yarovization Lysenko (water)	209	6.5	8.0	39
Hormonization with urine 75% + yarovization	198	6.3	8.3	59
Hormionization with urine 50% + yarovzation	242	7.8	10.4	55
Hormonization with urine 25% + yarovization	214	7.3	8.7	40

Table 2

Field plots	Hormonization of the seeds of wheat by soaking			The same + yarovization		
	Yield in g			Yield in g		
	Grain	Straw	Increase in grain in %	Grain	Straw	Increase in grain in %
Control (sowing with dry seeds)	59.1	69.9	—	—	—	—
Soaking in water	66.2	70.2	+12	50.0	66.7	—
Hormonization with urine 75%	72.0	79.5	+22	69.5	67.2	+17
Hormonization with urine 50%	65.4	70.7	+11	68.2	75.1	+15

Table 3

Field plots	Hormonization of seeds by soaking + yarovization		
	Grain in kg	Straw in kg	Increase in grain in %
Control (sowing with dry seeds)	4.11	8.73	—
Yarovization Lysenko (water)	4.40	7.89	+ 7.1
Hormonization with urine 50% + yarovization	5.46	10.32	+32.8
Hormonization with urine 25% + yarovization	4.48	9.80	+ 9.0

As is seen, almost all experiments showed a favourable effect of hormonization with certain concentrations of the urine. In isolated cases the increase in the yield reached 20—30%. The relation of grain to straw in hormonized sowings tended to be in favour of the former. A much greater effect was obtained by using the urine of non-pregnant mares.

Table 4

Field plots	Yield from 1 sq. meter plots in g		
	Grain	Straw	Increase in grain in %
Control sowing with dry seeds	56.1	56.9	—
Yarovization Lysenko water	56.0	56.7	—
Hormonization with urine 50% + yarovization	88.7	94.4	+30
Hormonization with urine 25% + yarovization	94.7	101.7	+60
Hormonization with urine 10% + yarovization	78.2	91.1	+32

Based on these data, as well as on extensive works conducted on artificial folliculine, which have not brought any satisfactory results, the conclusion is arrived at that the favourable effect of hormonization of seeds with urine is attributable to the action of vegetable hormones: auxins, bios, and other substances of Group B.

In experiments on hormonization of the seeds of wheat with yeast-extracts (race 12) the following results were obtained:

Table 5

Field plots	Hormonization by soaking			The same + yarovization		
	Grain in g	Straw in g	Increase in grain in %	Grain in g	Straw in g	Increase in grain in %
Control (sowing with dry seeds)	59.1	69.9	—	59.1	69.9	—
Yarovization Lysenko (water)	—	—	—	56.0	56.7	—
Hormonization with yeast 1:10	78.8	92.0	+33	74.1	86.5	+25
Hormonization with yeast 1:20	76.4	104.6	+29	78.0	94.1	+32

It is to be noted that while hormonization of the seeds with solutions of urine was facilitated by the lower temperatures required by the conditions of yarovization, in the case of yeast-extracts the effect was produced by a mere soaking of the seeds.

Unlike wheat, the other plant dealt with in our experiment—the rye—was entirely unresponsive to hormonization. The scope of this paper does not

permit of stating our assumptions on this subject. We shall only point out that we think that the conditions of hormonization of the seeds of different agricultural crops will greatly vary not only in respect of the concentration of the solutions applied, but also in regard to temperature conditions and duration of soaking; furthermore, the different structure of seeds of bean plants will, apparently, require a somewhat different approach to hormonization.

Beet was responsive to hormonization and its favourable effect began to manifest itself as early as from the first day of vegetation. As the experiment proceeded, the great difference in the leaves levelled down, but the difference in the weight of the root still remained. For reasons beyond our control we were, to our regret, unable to carry on our experiment till the end of the vegetation period, and had to remove the beet before due time, after a weight of 100—150 g had been reached by the root. The data relative to this experiment are tabulated in the following table:

Table 6

Field plots	Weight of root in g	Average weight of the root in g	Sugar %	Weight of the leaves in g
Control	70—101	88	17.1	108—149
Yarovization (water)	70—92	81	16.2	100—124
Hormonization with yeast 1 : 3 .	78—127	107	17.1	110—161
Hormonization with urine 50% .	122—151	134	17.1	162—213

The experimental work carried out by us in 1936 leads us to the conclusion that the practical possibility of utilizing hormonization for increasing the yield of agricultural crops is proved. A more detailed report on the results of the investigations of 1936 will be given in the journal «Biochemistry».

Laboratory of Biochemistry of the
All-Union Scientific Research
Institute of Northern Grain Farming
and Grain and Bean Crops.
Moscow.

Received
31. III. 1937.

REFERENCES

[1] N. G. Cholodny, Ber. d. D. bot. Ges., **42** (1924); Jahrb. wiss. Bot., **65** (1926); Biol. Zbl., **47** (1927); Planta, **21**, 4 (1934), **23** (1935). [2] Н. Г. Холодный, Природа, 8/9 (1933); 4 (1935); 3(1936); Сов. бот., 2 (1935); Усп. совр. биол., IV, 6 (1935); C. R. Acad. Sci. URSS. III, 8 (1936); III, 9 (1936). [3] V. Hartelius, Bioch. ZS., **261** (1933). [4] M. Janot, C. R., **198**. 1175 (1934). [5] Fr. Kögl, ZS. physiol. Chem., **214** (1933); **228** (1934); **235** (1935); **242** (1936); **243** (1936); Ber. d. D. Chem. Ges., **68** (1935). [6] Koschara, ZS. physiol. Chem., **232** (1936). [7] B. Nath, Nature, **134**, 27 (1934). [8] N. Nielsen, Planta, **6** (1928); Jahrb. wiss. Bot., **73** (1930), Bioch. ZS., **237** (1931); **249** (1932); **256** (1932). [9] R. Pohl, Planta, **24**, 523 (1935). [10] M. H. Roscoe, Bioch. Journ, 6, 1053 (1936). [11] Schander, ZS. f. Bot., **27** (1934). [12] В. Н. Товарницкий, Бюлл. ВАСХНИЛ, 6 (1936). [13] A. Virtanen, Nature, **132**, 408 (1932); Bioch. ZS., **272**, 32 (1934).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

PLANT PHYSIOLOGY

THE PROBLEM OF YAROVIZATION OF PERILLA

By B. S. ZAKHAROV

(Communicated by A. A. Richter, Member of the Academy, 11. IV. 1937)

The object of this work was to ascertain the possibilities of governing the development of perilla with the aid of yarovization.

Experiments were first carried out in 1935 with three specimens of Far Eastern perilla. Conditions of experiments: temperature, 15, 20 and 25°, moisture 40 per cent of weight of seed in air-dry condition, duration of yarovization 10 days. Sowing was done on plots of 3—4 m². Results of the experiments are set forth in Table 1.

Table 1

Number of specimen	Yarovization temperature (degrees) C	Acceleration of flowering compared with control (in days)
217	25	17
217	20	13
217	15	9
403	25	19
403	20	4
403	15	1
482	25	7
482	20	1
482	15	0

As is evidenced by the table, at a temperature of 25° flowering sets in in all the specimens at a noticeably earlier date than in the controls, i.e. at this temperature the perilla passes through the yarovization stages at a rapid pace. Yarovization proceeds at a temperature of 20 and 15° in specimen No. 217 but decidedly slower than at a temperature of 25°.

At temperatures of 20 and 15° yarovization in numbers 403 and 482 either does not proceed at all or occurs at a very slow pace. Owing to the fact that the plants suffered from drought the actual moment of maturity could not be fixed with any exactness.

Experiments on perilla from the Far East were repeated in the Winter of 1936 by the vegetational method in greenhouses. Yarovization was effected at a temperature of 25°, moisture 40 per cent, duration of yarovization 10 days. As a result of these experiments the formation of clusters, flowering

and maturity in the plants subjected to yarovization occurred 12 days sooner than in the control plants.

Experiments on the yarovization of Far Eastern sorts were carried out under field conditions in the Spring of 1936. Yarovization was conducted at a moisture of 40 per cent at duration and temperatures given in Table 2.

Table 2

Yarovization conditions		Acceleration of cluster formation compared with controls (in days)
Temperature (degrees) C	Duration (days)	
25	16	17
25	13	12
25	11	9
25	8	7
25	6	5
25	4	3
15	11	4
20	11	5
30	4	5
35	4	12
40	2	4
40	1	7
40	0.5	4
45	2	4
45	1	4
45	0.5	6

Owing to the prevalent climatic conditions, a second vegetative growth appeared following the formation of clusters, which retarded the onset of flowering and levelled the individual differences at the moment of maturity. Conclusions from these experiments may therefore be drawn only on the basis of data pertaining to the acceleration of formation of clusters (Table 2).

The data referred to indicates that the most superior variant of yarovization is at a temperature of 25°, duration 16 days, moreover, the longer the duration of yarovization the sooner dose formation of clusters set in. Though variants at a temperature of over 25° result in the accelerated formation of clusters, they incur the risk of hypergrowth or desiccation of the seed and are therefore not to be recommended.

Yarovization for an 11 day period at a temperature of 25, 20 and 15° produced the quicker formation of clusters by 8, 4 and 3 days, respectively. This confirms our conclusions of 1935 in regards to these temperatures.

We do not quote the crop yield as same were not obtained in the first experiment and in the second and third experiments same so small that no conclusions could be drown on the basis of the figures obtained.

A great shortcoming in the yarovisation of perilla was the infection of the plants with fungi. In 1936 V. P. Yagodkina, the phytopathologist of the Institute of Oil Bearing Cultures, conducted experiments in combatting fungi during the yarovization of perilla, with beneficial results.

The test of the effectiveness of yarovization as an agricultural process is to be conducted on a geographic aspect during 1937 by the Institute of Oil Bearing Cultures. During the current year experiments will commence on the yarovization of perilla at different periods.

Summary

1. Perilla reacts to yarovization, speeding up the formation of reproductive organe.

2. Superior conditions for yarovization are represented by a temperature of 25°, moisture 40 per cent and duration 10—16 days.

3. Yarovization proceeds at a very slow pace at temperatures below 25°. At temperatures exceeding 25° yarovization proceeds rapidly but is frought with the risk of desiccation or hypergrowth of the seed.

Laboratory of Agrophysiology.
Institute of Oil Bearing Cultures.
Krasnodar.

Received
11. IV. 1937.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

PLANT PHYSIOLOGY

THE EFFECT OF THE OSMOTIC PRESSURE OF SOIL SOLUTION ON THE YIELD AND COMPOSITION OF SUGAR-BEET

By T. T. DEMIDENKO and N. P. MARTYNOV

(Communicated by A. A. Richter, Member of the Academy, 26. III. 1937)

A number of workers have proved that cultivated plants both develop better and absorb mineral substances better when the nutrient medium contains a weak concentration of salts, inasmuch as, in this case, the latter disintegrate to a greater extent into ions and are distinguished by greater mobility, owing to which the electrolytes utilized by the plant for synthesis are easier absorbed and imbibed into the cells of the absorbent part of the root.

Although the sugar-beet originates from relatives that grow on the coasts of the Mediterranean and are distinguished by their high salt-resistance and also by their capacity of accumulating large amounts of mineral substances, the existing cultivated varieties of this plant painfully respond, when young, to the introduction of high doses of fertilizers, while at later stages of growth they stand very high concentrations of salt in the soil and continue to develop fairly well. Under agricultural conditions, the soil solution is the plant's chief source of mineral nutrient substances. The concentration of this solution depends on the moisture-content of the soil, on the total quantity of salts and on their solubility. If the humidity of the soil is high enough, the beet grows well notwithstanding the introduction of very high doses of fertilizers. But if the moisture-content drops, the plants immediately show signs of depression.

As the task of obtaining high crops of beet is inevitably associated with the introduction of high doses of fertilizers, research workers are confronted with the problem of determining the concentrations or degrees of osmotic pressure of the soil solution at which the sugar-beet will give the highest yield, both as regards quantity and quality.

On the other hand, it is most important to study the response of the sugar-beet to high doses of the various salts, by raising the osmotic pressure, and then to determine the dependence of the salt-resistance of the cultivated beet on the properties of the nutrient substances.

By introducing into the soil, in addition to a balanced mineral fertilizer, increased doses of $NaCl_3$, KCl or $CaCl_2$ having a common anion Cl' but different cations, and by recording the osmotic pressures produced by them in the soil, we purposed to study, not only the effect of osmotic pressure of the soil solution, but also that of each of the cations studied, on the yield and composition of the beet-root. It was also important to compare the

physiological value of the Cl′ and SO_4′ anions entering into the composition of «raw» potassium salts, and to establish their influence on the qualitative composition of the beet.

Lastly, in a series of variants, different degrees of osmotic pressure were produced by balanced nutrient solutions obtained by the introduction of $NaNO_3$, $Ca(H_2PO_4)_2$ and KCl in amounts necessary to set up definite osmotic pressure in the soil, these amounts being calculated after van't Hoff's formula.

For the purpose of solving the above-stated problems we organized in 1935 an experiment for investigating, under laboratory conditions, the influence of the osmotic pressure of soil solution on the sugar-beet.

The experiment was made on the background of a balanced fertilizer, N being introduced as $NaNO_3$, P_2O_5 as $Ca(H_2PO_4)_2$, and K_2O as KCl; 0.2 g of the active principle were introduced per kilogram of soil. The vessels contained 15 kg of absolutely dry soil each. B, Mn and Cu were introduced into all the vessels to the amount of 45 mg of each, i. e. 0.3 mg of each substance per kilogram of soil. These elements were added to force the growth of the plants and to increase the sugar content of the beet-root. The water-holding capacity of the soil was 42.57%, the soil being watered up to 60 per cent by weight of its water-holding capacity.

Date of sowing, May 9; seedlings appeared on May 15; last thinning out made of June 3.

Phenological observations of the beet cultivated at different osmotic pressures showed the following:

1. The lower was the concentration of the salt which produced the osmotic pressure of the soil solution, the better was the growth of the beet.

2. During the first half of its vegetative life, till the development of the foliage, the beet painfully responds to high salt-content of the soil, while in the second half of summer it easily endures high salt concentrations.

3. At a pressure of one atmosphere, soduim chloride and sulphate stimulate the growth of the young beet, while all the other salts did not produce any positive effect on the growth of the young beet.

4. The difference in the effect of chlorides on the growth of beet, independently of the magnitude of osmotic pressure, appears to be due to two factors: a) the different specific effects of the various cations on the metabolism in the plant, and b) the different absolute content of the various salts introduced for the purpose of producing different degrees of osmotic pressure of the soil solution.

In the following table, we give the data on the influence of osmotic pressure of the soil solution on the yield and composition of the sugar beet.

The above data justify the following conclusions.

1. Though Na cannot functionnally replace potassium in the organism of the sugar-beet, it is nevertheless absorbed by this plant, taking an active part in the distribution of the nutrient substances in the vegetative organs and enabling them to economise other mineral substances.

2. Of the chlorides included in the experiment, NaCl produces, in low concentrations, the most favourable influence on the growth of the young beet.

3. At its initial stages of growth, the beet suffers considerable depression in its development if osmotic pressure is high, but it recovers during the latter half of summer and continues to develop normally.

4. By producing different degrees of osmotic pressure in the soil through the agency of various salts, we succeeded in arranging the salts in the following series according to their physiological value: $NPK > KCl > NaCl \geqslant Na_2SO_4 > CaCl_2$.

Influence of Osmotic Pressure of the Soil Solution on the Yield and Composition of the Beet

№	Experimental scheme	Yield		Percentage of sugar	Colloids in 100 cm³ of diff. juice in g	Specific gravity of the juice	Viscosity of the juice	Osmotic pressure in atmospheres	Transpiration coefficient	Percentage of total nitrogen	Ash content in the root, per cent	Amount of K_2O absorbed, per cent	Amount of Na_2O absorbed, per cent
		grams	per cent										
1	Control (NPK)	515.41	100.0	18.87	0.2963	1.0961	2.4112	22.34	301.5	0.7056	2.6785	0.4578	0.0997
2	Soil + NPK + NaCl (1 atm) . .	556.75	108.0	20.47	0.4062	1.0984	2.4103	23.16	307.6	0.7604	2.9478	0.7021	0.1368
3	Soil + NPK + 3 NaCl (3 atm) .	609.10	118.0	19.30	0.5845	1.0995	2.5018	24.19	318.4	0.8605	3.5479	0.4328	0.1836
4	Soil + NPK + 5 NaCl (5 atm) .	555.91	108.0	19.17	0.6224	1.0986	2.5639	24.78	325.3	0.9196	3.9348	0.3451	0.2437
5	Soil + NPK + Na_2SO_4 (1 atm) .	593.63	115.0	19.36	0.3435	1.0985	2.3981	23.17	317.4	0.6144	3.2365	0.4346	0.0997
6	Soil + NPK + 3 Na_2SO_4 (3 atm)	622.53	120.0	18.75	0.4254	1.1034	2.5763	24.16	305.5	0.7235	3.4472	0.4018	0.0763
7	Soil + NPK + 5 Na_2SO_4 (5 atm)	556.90	108.0	18.42	0.5736	1.0965	2.6689	25.17	285.6	0.8745	3.9876	0.3647	0.0445
8	Soil + NPK + KCl (1 atm) . .	610.72	122.0	20.80	0.3741	1.0984	2.4521	22.78	304.3	0.8574	3.0689	0.8193	0.1145
9	Soil + NPK + 3 KCl (3 atm) .	648.56	132.0	19.50	0.4583	1.0885	2.4680	23.18	316.8	0.7905	3.2768	0.8296	0.0836
10	Soil + NPK + 5 KCl (5 atm) .	548.32	110.0	19.17	0.6785	1.0896	2.4723	24.15	325.5	0.9113	3.6787	0.8836	0.0748
11	Soil + NPK + $CaCl_2$ (1 atm) .	609.63	118.0	18.52	0.3642	1.0887	2.5376	24.18	303.4	0.6836	3.0786	0.7814	0.1149
12	Soil + NPK + 3 $CaCl_2$ (3 atm) .	514.54	97.0	17.22	0.4815	1.0894	2.6432	25.06	318.6	0.7576	3.6682	0.6496	0.1546
13	Soil + NPK + 5 $CaCl_2$ (5 atm) .	304.54	67.0	17.13	0.5236	1.0899	2.6846	25.98	324.5	0.9114	3.9786	0.5238	0.2037
14	Soil + $NaNO_3$ + KCl + + $Ca(H_2PO_4)_2$ (1 atm)	685.36	137.0	20.85	0.3954	1.0932	2.3578	23.17	306.5	0.6436	3.1126	0.6536	0.0672
15	Soil + 3 [$NaNO_3$ + KCl + + $Ca(H_2PO_4)_2$] (3 atm) . . .	754.78	147.0	20.13	0.4283	1.0946	2.4335	24.86	311.2	0.6954	3.5473	0.4239	0.0984
16	Soil + 5 [$NaNO_3$ + KCl + + $Ca(H_2PO_4)_2$] (5 atm) . . .	554.12	111.0	19.24	0.4725	1.0972	2.5786	25.72	318.4	0.8934	3.8641	0.3918	0.1045

5. The curve characterizing the dependence of the yield of beet on the concentration of the salts introduced into the soil shows the maximum yield to occur at pressures of 1—3 atmospheres, while in certain cases a pressure of 5 atmospheres produces a depression in the growth and yield of beet-root.

6. According to the yield of beet the salts in question form the following series: NPK>KCl>NaCl≥Na_2SO_4>$CaCl_2$.

7. The higher the concentration of the soil solution the greater the amount of water transpired by the beet during the vegetative period. According to the amount of moisture evaporated by the plants under the effect of the various salts, the latter may be arranged in the following series: $CaCl_2$> NaCl>KCl.

8. Increasing doses of salts introduced into the soil for the production of different degrees of osmotic pressure considerably affect the accumulation of hydrophylous colloids. The action of the different cations on the formation of colloids may be shown by the following series: NaCl>KCl>Na_2SO_4> NPK>$CaCl_2$.

9. The nitrogen content in the root increases with the rise of osmotic pressure of the soil solution, with the exception of plants that receive increasing doses of potassium, where no such correlation between the nitrogen and osmotic pressure has been observed.

10. The beet absorbs different amounts of potassium depending on the magnitude of osmotic pressure produced by the various salts. The gradation in the amount of potassium absorbed by the beet during the vegetative period under the action of the various salts may be shown by the following series: KCl>NaCl>Na_2SO_4>NPK>$CaCl_2$.

11. Alongside of potassium, sodium was also absorbed by the beet. The greatest intake of sodium was observed to occur at increasing doses of NaCl and Na_2SO_4, while the intake of sodium in the vessels with high doses of KCl was considerably less.

12. With increasing osmotic pressure of the soil solution the mineral content of the beet also increases under the action of the salts. The increase of the mineral content under the action of the various salts may be shown by the following series: KCl>NaCl>Na_2SO >NPK>$CaCl_2$.

13. With increasing concentration of the salts, the osmotic pressure of the cell juice also increases, this increase being not slower than that of the amount of the salt introduced.

14. The specific gravity of the juice is not affected by increasing osmotic pressure, but its viscosity and surface tension increase under the influence of the different amounts of the salts introduced into the soil and of their dispersive or coagulative action.

15. Considering the high osmotic pressure of the soil solution and of the cell juice of the beet, as well as its mineral content, it is necessary to emphasize the fact that the beet plant is very salt-resistant, especially after the development of its foliage and till the end of the vegetative period, while during the initial stages of growth it responds painfully to high salt content of the soil solution. Therefore, only small doses of fertilizer should be supplied to the beet at the beginning of the vegetative period, while later on the doses should be increased with growth.

Department of Physiology,
Voronezh Experimental Station.

Received
26. III. 1937.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

PLANT PHYSIOLOGY

THE INFLUENCE OF LIGHT ON THE INFLOW OF NUTRIENT SUBSTANCES IN PLANTS

By T. T. DEMIDENKO and V. P. GOLLE

(Communicated by A. A. Richter, Member of the Academy, 16. IV. 1937)

The study of the absorption of nutrient substances by plants has been conducted by many authors ([1–9]). As a result of these studies it was found that the acquisition of ash elements and nitrogen occurs fairly evenly in certain plants (potatoes, beet) whilst im others (cereals) three periods are to be noted.

In the first period (5—6 weeks following germination) the plant manifests its peak point of activity in absorption of nutrient substances; the second period, the transition from the vegetative to the reproductive phase, witnesses a decrease in absorption of electrolytes, and, in some cases, complete cessation and even their discharge. Absorption is once again witnessed at the close of the period of grain formation. The third, period of maturity, is accompanied by complete cessation of the inflow of nutrient substances and the partial reduction of their general content in the plant in connection with the falling off of leaves.

In studying the daily course of NPK absorption by wheat shoots and in placing them for definite periods in nutrive solutions, Breazeale ([10]) found that during the first two hours in the solution wheat absorbs potassium to the extent of 33.7 per cent of the total quantity absorbed by the plants in 24 hours.

Studies by Niclewcky, Krause and Lemanczyk ([11]) regarding the absorption of potassium from salts of KCl during various periods of time found that the flow of potassium in the plant was irregular and ceased completely during the night time.

In addition to this data we also have the results of the study of the influence of light on root nutrition.

In studying the daily course of the flow of potassium, nitrogen, phosphorus and calcium, Domontovich ([12]) drew the conclusion of the specific nature of flow of each element, potassium, in particular, being absorbed by the plant to a greater extent in the first half of the day compared with the second half.

Cachioni-Walter, in the study ([14]) of the problem of daily consumption of nutrients by plants, ascertained that oats when steeped in a nutrive solution

Table 1

Effect of Duration of Working (Light) Day on the Rate of Maturity of Plants

№	Culture	Kind	Length of working day (hours)	Date of sowing	Sprouting	Clustering	Ears formation	Flowering	Maturing	Harvesting	Duration of vegetative period (days)
1	Spring wheat . .	Erithrospermum 0841	8		16 V	4 VI	—	—	—	9 VIII	—
			12	13 V	17 V	2 VI	2 VIII	—	—	9 VIII	—
			Full day		17 V	2 VI	26 VI	3 VII	4 VIII	9 VIII	73
2	Oats	Pobeda	8		18 V	6 VI	—	—	—	9 VIII	—
			12	13 V	18 V	4 VI	7 VII	—	—	9 VIII	—
			Full day		18 V	4 VI	26 VI	5 VII	5 VIII	9 VIII	77
3	Mustard	Sizaya Sarepta	8		15 V			—	—	9 VIII	—
			12	13 V	15 V	—	—	20 VII	—	9 VIII	—
			Full day		15 V			20 VII	3 VIII	9 VIII	78
4	Maize	Minnesota	8		18 V		21 VI	14 VII	22 VIII	22 VIII	94
			12	13 V	18 V	—	27 VI	9 VII	19 VIII	22 VIII	92
			Full day		18 V		7 VII	14 VIII	—	22 VIII	—
5	Millet	White 198 СХИ	8		20 V	14 VI	16 VI	—	20 VII	25 VII	60
			12	13 V	20 V	10 VI	16 VI	—	23 VII	25 VII	63
			Full day		20 V	10 VI	16 VII	25 VII	23 VIII	25 VII	93
6	Buckwheat . . .	Bogatyr	8		19 V			14 VI	3 VIII	6 VIII	74
			12	13 V	19 V	—	—	11 VI	4 VIII	6 VIII	75
			Full day		19 V			16 VI	25 VIII	26 IX	96
7	Soya	Krushulja 9/3	8		18 V			27 VI	4 VIII	5 VIII	76
			12	—	19 V	—	—	25 VI	4 VIII	5 VIII	76
			Full day		19 V			10 VII	25 VIII	27 VIII	96

Table 2

Crop and Results of Chemical Analysis

№	Culture	Working day (hours)	Crop		Nitrogen		Phosphorus		Potassium		Calcium		Discharge			
			Total	Grain	Grain	Straw	Grain	Straw	Grain	Straw	Grain	Straw	N	P_2O_5	K_2O	CaO
1	Spring wheat . .	8	15.25	—	—	3.9548	—	0.8135	—	0.5010	—	0.1329	0.6031	0.1241	0.0764	0.0202
		12	23.80	—	—	3.4930	—	0.4830	—	0.4615	—	0.0858	0.8313	0.1149	0.1098	0.0204
		Full day	26.40	10.7	2.6943	0.7732	0.9916	0.3622	0.3130	0.3818	0.0394	0.0754	0.4097	0.1629	0.0856	0.0160
2	Oats	8	33.5	—	—	3.4930	—	0.7992	—	0.7334	—	0.1198	1.1172	0.2681	0.2461	0.0402
		12	50.3	—	—	3.3896	—	0.5825	—	0.5840	—	0.0850	1.7067	0.2933	0.2940	0.0428
		Full day	57.9	27.5	2.1509	0.8639	0.6442	0.1235	0.2530	0.4728	0.0486	0.0618	0.8541	0.2097	0.2133	0.0267
3	Mustard	8	39.4	—	—	1.7365	—	0.7181	—	0.3320	—	0.2980	0.6842	0.2829	0.1308	0.1174
		12	73.8	—	—	0.4074	—	0.4875	—	0.3524	—	0.1904	0.3009	0.3600	0.2602	0.1406
		Full day	84.6	14.0	3.2169	0.4609	1.8234	0.1495	0.2932	0.3600	0.1922	0.1768	0.7756	0.3608	0.2950	0.1517
4	Maize	8	44.3	8.5	1.4764	0.9671	0.7864	0.3681	0.5917	0.3879	0.0041	0.0634	0.4716	0.1986	0.1891	0.0230
		12	74.0	16.0	0.9994	1.0629	0.7011	0.2184	0.3541	0.3476	0.0039	0.0653	0.7806	0.2397	0.2596	0.0388
		Full day	80.2	—	—	1.1517	—	0.1497	—	0.3434	—	0.0779	0.9237	0.1200	0.2754	0.0625
5	Millet	8	3.9	2.2	1.5574	1.5880	1.1351	0.6090	0.4036	0.6270	0.0811	0.1856	0.0620	0.0359	0.0196	0.0040
		12	16.3	8.8	2.4333	1.0755	0.6972	0.2219	0.3361	0.5989	0.0734	0.2191	0.2950	0.0780	0.0748	0.0230
		Full day	47.2	11.5	3.4329	1.8641	0.5445	0.0998	0.2703	0.5199	0.0562	0.1414	1.0610	0.0983	0.2167	0.0570
6	Buckwheat . .	8	9.8	2.1	1.1481	1.7056	1.0740	1.2616	0.3360	0.3804	0.3606	0.8459	0.1547	0.1193	0.0362	0.0717
		12	23.7	7.3	1.2579	1.9290	0.9768	0.7496	0.2750	0.3930	0.3623	0.6144	0.4091	0.1946	0.0847	0.1276
		Full day	58.7	32.0	2.0530	2.3609	0.8282	0.2068	0.3821	0.4013	0.3764	0.3507	1.2870	0.3203	0.2294	0.1240
7	Soya	8	17.3	7.0	6.9759	1.8714	2.0516	1.0814	0.6640	0.3054	0.0605	0.4347	0.8106	0.2553	0.0785	0.0491
		12	23.8	7.5	7.5784	0.8564	1.7736	0.8223	0.8029	0.3264	0.0231	0.3339	1.3792	0.2674	0.1136	0.0563
		Full day	55.4	15.9	4.5351	2.2629	1.1902	0.5606	0.4409	0.3485	0.0182	0.2123	1.6149	0.4106	0.2076	0.0865

absorb potassium at a more intensive rate during the first three hours in such solution and that the age of the plant yields an influence on the distribution of potassium in the plant organism.

Phosphorus is absorbed mainly during the second half of the day. Absorption of potassium and phosphorus during the night time is very small. Cachioni-Walter did not note the discharge of phosphorus by plants into the outer surroundings during her experiments.

Akhromeiko, in his studies ([13]) noted the discharge of phosphoric acid by the plants. Upon surplus absorption during the day when an intensive inflow occurs, the phosphoric acid, not bound with the products of assimilation, together with the downflowing currents is borne along the phloemic section of the fibrous-vessel system from the leaves to the roots and is exosmosed into the outer surroundings.

Conclusive evidence of this assumption is found in experiments conducted with the isolated nutrition of a number of plants. In these experiments, the plants, not deriving phosphorus from one of the inner vessels, developed fairly favourably on account of the phosphoric acid which was exosmosed from the cell juices into the inner vessel and was utilized by the plant for nutrition.

The above experiments indicate that light nutrition exerts much influence on the inflow of nutrient substances into the plant.

The object of this study was to follow the flow of nutrient substances in plants of long and short days, developed under different periods of working time of light; to determine whether the absorption of nutrient elements occurs at a similar or varying rate in these plants.

We have studied the course of the flow of nutrient elements in plants under long and short days, cultivating them under 8 hours, 12 hours and full day, taking tests every fifth day, independent of the stage of development of the plants. We give here the results fixing the flow of nutrient substances in long and short day plants, reared under 8 hours, 12 hours and full day and which were removed on attaining maturity. In the next communication we shall give a description of the course of flow of ash elements in the long and short day plants for every fifth day and also according to the phases of their development.

The experiment was conducted under vegetative conditions. The long day plants which were studied were spring wheat, oats, mustard, the short day plants being maize, millet, buckwheat and soya.

Data of the phenological studies are given in tabl. 1.

Full day plants developed normally and completed their cycle of development.

In short day plants with an increased light regime the vegetative period was lengthened, a reduction of the regime resulting in contraction of this period, the alteration in the duration of vegetative period, moreover, did not bear a uniform nature. Maize, for instance, matured only under 8-hour and 12-hour days, whereas under full day, maturity did not set in even following frosty weather. Buckwheat, millet and soya resulted in a fairly pronounced reaction to the duration of daylight, sharply protracting the period of maturity under a full day.

The crop yield and results of chemical analysis are set forth in Table 2.

The following is evident from the above data:

With the increase of working light time long and short day plants give a sharply increased yeild. The grain yield in a long day plant was obtained only under a full working day. With the exception of maize the short day plants under a full day yielded a grain crop throughout all the test variants,

the yield sharply increasing with the duration of daylight reaching its maximum under full day.

In regards to the absorption of nutrient substances by long and short day plants reared under various periods of daylight the following picture was noted: the comparative content of nitrogen in the straw decreases with the lengthening of the day in wheat, oats and mustard, whilst an increase was noted in maize and buckwheat. The greatest content in millet and soya was found under a full day, the smallest under a 12-hour day.

The comparative content of nitrogen in the grain increases with the lengthening of the working day in millet and buckwheat, but decreases in the case of maize. In soya grains the highest content was found to be under a 12-hour day, the lowest under a full day. The discharge of nitrogen in long day plants increases under a 12-hour day and falls under a full day. The maximum in mustard is reached under a full day, the minimum under a 12-hour day.

The discharge of nitrogen in short day plants increases with the lengthening of the day. The comparative content of phosphorus in the straw and grain diminishes with the increased working day in both groups of plants.

Discharge of phosphorus increases in mustard, millet, buckwheat and soya following the lengthening of the working day. The greatest discharge of phosphorus in spring wheat is under a full day and in oats and maize under a 12-hour day.

The comparative content of potassium diminishes with the lengthening of the day in spring wheat, oats, maize and millet, increasing in buckwheat, mustard and soya. Potassium content in the grain diminishes in maize and millet. The maximum in buckwheat is attained under a full day, the minimum under a 12-hour day, the maximum in soya under a 12-hour day and the minimum under a full day.

With the lengthening of the day, the discharge of potassium is increased in maize, millet, buckwheat, mustard and soya. The maximum in spring wheat is attained under a 12-hour day, discharges under an 8-hour and full day being very similar in quantity. The maximum in oats is obtained under a 12-hour day, the minimum under a full day.

With the increase of the length of the daylight the calcium content in the straw is decreased in spring wheat, oats, buckwheat, mustard, soya and millet, whilst showing an increase in maize.

Calcium content in the grain, following the lengthening of the day, diminishes in maize, millet, soya and increases in buckwheat.

Discharge of calcium is increased following the lengthening of the day in maize, millet, mustard and soya. The maximum discharge in buckwheat, oats and spring wheat is reached under a 12-hour day, the minimum under a full day.

Summary

1. Alteration in the length of daylight causes profound changes in the development of plants.

2. Long day plants, when transferred to short day conditions (8 and 12 hours), form massive vegetative growth.

3. Short day plants react to changes in the duration of the working (light) day by corresponding alterations in the duration of the vegetative period. Upon lengthening the day the vegetative period is substantially increased.

4. The crop yield increases in direct proportion to the length of the day.

5. The discharge of nutrient elements by short day plants increases with the lengthening of the working (light) day which indicates the direct dependence on photosynthesis of the inflow of nutrient elements into the plant.

6. There are insufficient foundations for the deduction of any conclusion regarding the dependence of the discharge of nutrient elements by long day plants, in connection with alteration of duration of the period of light, as the plants under short days did not reach the stage of maturity.

Department of Physiology.
Voronezh District Experimental Station.

Received
16. IV. 1937.

REFERENCES

[1] J. Adorjan, Journ. f. Landw., 50 (1902). [2] H. Wilforth, Romer, Hund, Wimmer, Landw. Versuchst. (1906). [3] W. J. Jones a. H. A. Huston, Ind. Agr. Exp. Stat. Bull. (1914). [4] J. S. Burd, Journ. of Agr. Res. (1919). [5] D. R. Hoagland, Journ. Agr. Res. (1919). [6] В. А. Чижов, Накопление сухого вещества, зольных и азотистых веществ в культурных растениях в полевых условиях. [7] O. Arrheims, Journ. Gen. Phys. (1923). [8] K. Pirschle, Planta (1929). [9] М. К. Домонтович, Тр. ВИУ, вып. 52 (1928). [10] F. Breazeale, Journ. of Agr. Res. (1923). [11] B. Niclewcky, A. Krause u. K. Lemanczyk, Jahrb. f. wiss. Bot. (1928). [12] М. К. Домонтович и А. И. Грошенков, ААЖ (1930). [13] А. И. Ахромейко, ИМЕН (1935). [14] Л. С. Качиони-Вальтер, Физиол. растений, вып. 8., ВИУАА (1935).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

EXPERIMENTAL ZOOLOGY

THE EFFECT OF THE SUBCUTANEOUS INSERTION OE EPITHELIAL, OSSEOUS AND MUSCULAR TISSUES ON THE SURROUNDING TISSUES IN AXOLOTLS

By N. V. NASSONOV, Member of the Academy

In my papers «The Rôle of Cartilage in Formations in Axolotls»(1) and «Influence of Various Factors on Morphogenesis Following Homotopical Subcutaneous Insertions of Cartilage in the Axolotl» (2). I assume that the cartilage, on being inserted subcutaneously, acts as an organizer, namely, it sets up, under definite conditions, a series of processes, which result in new formations. So, for instance, the subcutaneous insertion of cartilage under the skin of the leg may cause the development of an additional (duplicate) leg or part of a leg.

In the present paper, I purpose to determine the effect produced on the surrounding tissues (cartilage, muscle, etc.) by epithelial, osseous and muscular tissues inserted subcutaneously, and also whether they can act as organizers, similarly to cartilage. Pieces of all these tissues were taken from axolotls, about one year old, and were inserted under the skin of the leg.

The epithelium was taken from the surface of the blastema, formed after the amputation of a leg. Pieces of the epithelium, after being freed in Ringer's solution from the mesenchymal cells adhering to them, were inserted under the skin subcutaneously after the lapse of different periods of time after the amputation. The total number of implantations made was 125; 40 implantations were made 4 days after the amputation of the leg, 20 implantations were made after 6 days, 25 after 8—9 days, 20 after 11—12 days and 20 after 13 days.

No new formations were observed to occur for over half a year. There was only a small swelling and a reddening of the skin at the places of insertion.

To elucidate the processes occurring within, 15 series of sections were made through the place of insertion, namely, of the 4-day-old epithelium on the 6th (two series), 10th, 20th, 27th, 40th, 50th and 193rd day after its insertion; of the 6-day-old epithelium on the 13th and 93rd day; of the 8—9-day-old epithelium on the 13th, 23rd, 30th, 40th and 50th day; and of the 12-day-old epithelium on the 65th day after insertion. If the pieces of epithelium inserted into the leg were large (about 2—2.5 mm) they formed a fold or curled into a tube, turning their inner surface toward the surrounding tissues. On the 6th and 10th day (in 4-day-old epithelium), phagocytes begin to collect around the inserted piece of epithelium, and its incipient disintegration

may be observed. We find a similar picture on the 13th day (in 4- and 8-day-old epithelia), but no signs of the disintegration of the neighbouring tissues, including the cutis, are to be observed. A complete disintegration of the implant was observed on sections made on the 20th day (in 4-day-old epithelium), the cutis above the place of insertion being also completely destroyed by the phagocytes. A complete disintegration of the implant was also recorded in sections made on the 23rd, 30th and 40th day (in 8-day-old epithelium), the cutis above the place of insertion being only partly destroyed. Sometimes an incipient destruction of the skin epithelium by phagocytes may be observed in sections made 50 days after implantation. Sometimes an incipient disintegration of the cutis with the epithelium remaining untouched by the phagocytes was observed only much later, namely, on the 193rd day.

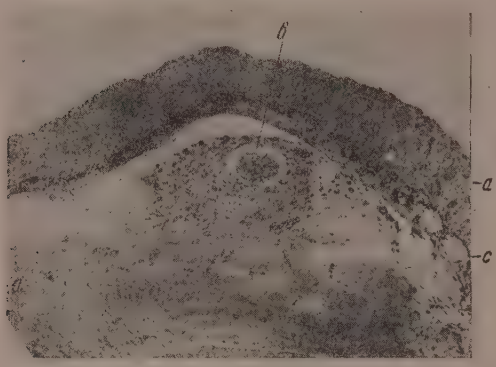


Fig. 1. Section on the 13th day after insertion under the skin of the 6-day-old epithelium from the blastema of the amputated leg: *a*—epithelium of the leg; *b*—the 6-day-old epithelium inserted under the skin; *c*—cutis; *d*—muscles of the leg.

Among the decomposition products may occur coriocide*, which assists in destroying the cutis.

Thus we see that, as long as the epithelium is not destroyed, the surrounding tissues are not exposed to any histolyzing action. This contradicts the opinion advanced by Efimov(3). In his paper entitled «The Rôle of the Skin in the Process of Regeneration in the Axolotl», he makes the suggestion (needing, as he says, further corroboration), that «the decomposition of tissues in the regenerative process occurs mainly... as a result of the interaction of the epithelial membrane and the wound surface». Efimov grounds his suggestion on his experiment of leaving a bit of skin under the skin with which the wound surface was covered after the amputation of the tail (page 363). However, both this experiment and the next are not convincing, mainly owing to the writer not having applied the method of serial sections for the purpose of determining the processes occurring under the skin. Besides, that on the basis of my experiments of subcutaneous insertions of the epithelium from a blastema, the development of a regenerate in Efimov's experiment may be explained as follows. The bit of skin left under the skin disintegrates, while the coriocide formed causes the destruction of the cutis, which probably hindered, as a component of the skin, the development of regenerative processes. It is very probable that, owing to the disintegration of the epithelium, a rupture of the skin occurred, as mentioned by the author, and thus conditions were set up for the origination of regenerative processes.

Efimov's other experiment (page 364), that of covering the wound surface with skin after the lapse of several days is also not conclusive, for from the standpoint of the results of my experiments the development of a regenerate after the wound surface of an amputated tail has already become

* See my above cited paper (2).

covered with epithelium, may be explained by the destruction of the cutis by the decomposition products of the epithelium lying under the skin. It is also possible that these decomposition products destroy the epithelium as well, and there occurs the rupture mentioned by Efimov.

Most probably, the disintegration of the epithelium caused also the disappearance of the cutis and the rupture of the skin when a 4-day-old blastema was covered by it in Orechowitch and Bromley's experiments ([4]) described by them in their paper entitled «The Histolyzing Properties of a Regeneration Blastema». Furthermore, as we know, under the blastema there can develop the effects of the cartilage, which, as has been shown in my paper «The Rôle of Cartilage in Formations in Axolotls», may contribute to the disintegration of the cutis and the development of regenerative processes ([1]).

Polejajev ([5],[6]) in his paper on the rôle of the epithelium in the initial stages of development of the regenerating bud of a limb also considers the «rôle of the epithelium to consist in the destruction and histolysis of the underlying tissues of the amputation surface» [([6]), p. 726].

In view of all the above-described, it seems to me that the available data do not justify the assumption that normal epithelium possesses histolyzing properties and that, acting on the tissue of the wound surface, it originates regenerative processes. The more so since the disintegration of the inserted tissue can occur when it is separated from the epithelium by the cutis, as, for example on the insertion of boiled muscle fibre, of cartilage heated to 90°, etc. I am inclined to believe that, in this case, the cause lies in the cutis and generally in the connective tissue with the collagen fibres, which probably inhibits the regenerative processes and that the histolysis of the tissue inserted under the skin is of a fermentative character and occurs without the participation of the epithelium.

In any case it is impossible to admit the presence of any organizing properties whatsoever in the epithelium. As some inductors manifest activity at the beginning of embryonic growth only in a dried state (Spemann, Holtfretter and others), I inserted subcutaneously 40 bits of dried epithelium. The results were negative, and no new formations were observed, even after the lapse of more than a year.

O s s e o u s t i s s u e, inserted under the skin, likewise produced no new formations. 60 insertions were made of this tissue taken from the outside bones of the cranium. 8 series of sections were then made on the 50th, 77th and 90th day. The bits of bone were incapsulated or were on their way outwards through the epithelium, they either lay wholly within it or had one end projecting outside. In one series, made after 77 days and in another series, made after 90 days, no inserted bits of bone were to be found. They seem to have been eliminated through the skin. At the places where the bits of bone had come out, the cutis was considerably modified and had the appearance of a thin layer of connective tissue.

50 subcutaneous insertions of dried bone were made. The same phenomena were observed as in the case of the insertion of normal bone. Two series of sections, made seven months later, showed that even very small bits of bone lie at different depths within the epithelium, as well as at its very surface. It therefore seems probable that all the insertions are eliminated as foreign bodies. 40 insertions of fine sand and infusorial earth were made under the skin of the legs and of the sides of the body; all these substances were eliminated.

M u s c u l a r t i s s u e was taken for subcutaneous insertions together with the connective tissue, which it is very difficult to separate from the mus-

cle fibres. A total of 56 insertions of parts of normal leg muscles was made. Observations carried on for over a year showed no new formations at the places of insertion. During a month or somewhat more an inflammation of the skin and small swellings were to be observed at these places. The inserted muscular tissue incapsulates, but the connective tissue of the capsule interrupts here and there over a small area, and the phagocytes penetrate into the capsule. Within the mass of the insertion cavities containing a liquid, probably the product of histolysis, sometimes develops. Sometimes during a long time no modifications whatsoever were recorded. Observations have shown the cutis to be intact even on the 134th day. In one case, a complete destruction of the implanted mass by phagocytes and the disappearance of the cutis were observed on the 46th day.

The picture is different if the muscular tissue subcutaneously inserted is preliminarily mechanically destroyed, for instance, crushed between glasses. In such cases, during the first days after the insertion semi-transparent white or pinkish vesicles or small abscesses may appear on the swelling over the implants; the skin cracks at such places and a liquid exudes.

Fig. 2. Section on the 3rd day after insertion under the skin of crushed muscular tissues: *a*—epithelium of the skin of the leg; *b*—crushed muscular tissues; *c*—the coagulated liquid under the epithelium; *d*—cutis; *e*—the failure of the epithelium.

58 insertions were made in all. For more than half a year no new formations were developed. Of the eleven series of sections, three were made three days after the subcutaneous insertion into the legs of one and the same animal. These sections showed that the cutis may be destroyed to greater or smaller extent. Phagocytes appear within the inserted mass and irregularly shaped cavities filled with liquid develop. In two cases, this liquid accumulated under the epithelium at those places where the cutis had been destroyed and had raised the epithelium. In one series of sections a rupture of the epithelium is observed at such a place. Sections made 7, 28, 34 and 38 days after the insertion showed different degrees of disintegration of the epithelium, but in one series of sections, that made on the 52nd day, I found the epithelium intact. The degree of disintegration of the epithelium probably depends on the degree of disintegration of the muscular tissue at the time of its insertion.

At the same time the cavities filled with liquid within the mass of the implant gradually diminish. They are not to be seen in sections made 275 days after the insertion and the growing activity of the phagocytes results in a complete destruction of the cutis and the whole mass of the insertion.

45 subcutaneous insertions were made with boiled and dried muscular tissue. Serial sections made 9 months later showed no signs of destruction of the cutis and no signs of new formations. Finally the cells of the insertion were destroyed by phagocytes.

Therefore, it appears that during the disintegration of muscle fibre inserted under the skin coriocide is produced, but no stimulation of epithelial proliferation occurs. The substances stimulating such a proliferation seem to be unable, in this case, to manifest their activity; though, as is known, proliferation of epithelial cells and its regulation occurs on the insertion of muscle fibre under the ectoderm at early stages of embryonic development [Woerdeman (7)]. It is to be supposed that, in the adult animal, the substances stimulating epithelial proliferation cannot manifest their activity as they are combined with other substances.

How this stimulator is set free, is unknown. Waddington, Needham and Brachet (8) believe that the organizer may experience the action of the substance (the activator) which breaks the bond mentioned above. In a live organizer, this breaking of the bond may occur, according to Needham, either following upon a change in the metabolism or as a result of a revercible denaturation of the proteins (p. 194), or according to the present author, the organizing substance may, under the influence of a changed surrounding medium, be set free as a result of an adsorptive expulsion from the micellar surfaces by some other, more strongly adsorbing substances. At the same time, the living organizers may, under the influence of a new surrounding medium (for instance by subcutaneous insertions) undergo decomposition, most probably of a fermentative nature, this being accompanied, similarly to what is observed in dead organizers, by the liberation of organizing substances.

Under the influence of a new environment, the muscle fibres probably undergo a change of the active substances stimulating proliferation of the epithelium being liberated and at the same time there arising regulators of this proliferation. These regulators cause the formation of a rudiment of the central nervous system. In his short paper, Woerdeman does not say whether there occur any structural changes of the muscular tissue after its insertion under the ectoderm and whether it disintegrates.

Moreover, in the case of a subectodermal insertion of muscle fibre, the longitudinal axis of the secondary rudiment of the nervous system follows the long axis of the fibre. Woerdeman says: «The graft, however, can determine the direction of the axis as I was able to demonstrate by using long-stretched bits of an isolated muscle fibre as inductor. The axis of the induced neural plates was determined by the longitudinal axis of the muscle fibre» (p. 312).

It has not yet been ascertained to what extent the ectoderm, on the one hand, and the organizing substance of the organizer, on the other, participate in the formation of the secondary rudiment of the nervous system. Different conjectures have been advanced in papers dealing with this problem [Woerdeman, 1936 (7), Waddington and Needham 1936 (9)].

Summing up all the above stated, we come to the following conclusions:

1. Subcutaneous insertions of epithelial, osseous and muscular tissue do not produce any new formations in adult axolotls. The same results are obtained when dry tissues are used.

2. Epithelium inserted under the skin whilst it is not yet disintegrated does not exert an histolytical effect on the adjoining tissues.

3. There is no reason to believe, together with Efimov, Poležajev, and others, that the epithelium of a blastema in a normal condition acts histolytically on the adjoining tissues during regenerative processes.

4. Osseous tissue, inserted under the skin, either becomes incapsulated or is eliminated through the epithelium.

5. During its disintegration under the skin, muscle fibre may cause disintegration of the cutis. Preliminary crushing of the muscle fibre between slides enhances its action on the cutis.

6. The substance which, as I believe, destroys the cutis and which I have named coriocide, appears probably only after the desintegration of the muscle fibres and is a product of their disintegration. High temperature seems to eliminate the action of coriocide or to modify it.

7. In adult axolotls, the organizing substances of the muscle tissue do not manifest any activity because they are probably combined with other substances in this tissue. As Woerdeman has shown, the insertion of muscle fibre under the ectoderm of an embryo causes the formation of the neural rudiment. This manifestation of the activity of the organizing substances may be tentatively explained by their liberation under the ectoderm owing to the action of the substances of the new environment. Biochemical investigations are needed to solve this problem.

N. V. Nassonov's Laboratory for the Study of the Problem of Organizers in the Animal Organism. Academy of Sciences of the USSR. Moscow.

Received 7. IV. 1937.

REFERENCES

[1] N. V. Nassonov, C. R. Acad. Sci. URSS, I, № 6, 414 (1935). [2] N. V. Nassonov, Ibid., XIII, № 2, 97 (1936). [3] И. М. Шумов, Журн. эксперим. биол., VII, вып. 3, 352 (1931). [4] W. N. Gretschowitsch a. N. W. Bromley, Biolog. Zentralblatt, 524 (1934). [5] И. В. Шумович, Зоол. журн., XV, 177 (1936). [6] L. W. Polejajev, Arch. f. Entw.-Mech., 133, 701 (1935). [7] M. Woerdeman, Proc. Akad. van Wedensch. Amsterdam, XXXIX, № 9, 806 (1936). [8] C. H. Waddington, J. Needham a. J. Brachet, Proc. Roy. Soc., London, Ser. B., № 17, 120, 179 (1936). [9] C. H. Waddington a. J. Needham, Proc. Akad. van Wedensch. Amsterdam, XXXIX, 7, 886 (1936).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

EXPERIMENTELLE ZOOLOGIE

ÜBER DIE DETERMINATION DES REGENERATS EINER EXTREMITÄT BEIM AXOLOTL

Von L. V. POLEŽAEV

(Vorgelegt von I. I. Schmalhausen, Mitglied der Academie, d. 3. II. 1937)

Wir bezeichnen eine regenerierende Extremitätsanlage, welche nicht zur selbständigen, herkunftsgemässen Entwicklung fähig ist, als undeterminiert, wenn diese Anlage von alten unter ihr liegenden Geweben des Organrestes abgetrennt ist und alle notwendigen Bedingungen zur Entwicklung besitzt. Eine determinierte Anlage hingegen ist eine solche, welche bei eben den gleichen Bedingungen die Fähigkeit zur selbständigen, herkunftsgemässen Entwicklung besitzt. Als Prozess der Determination des Regenerats bezeichnen wir den Prozess, welcher den Übergang der regenerierenden Anlage aus dem Stadium der Undeterminiertheit in das Stadium der Determiniertheit bestimmt.

Experimenteller Teil

In allen Versuchsserien wurde das Experiment gleichartig ausgeführt. Beim Axolotl wurde im Gebiete des Rückens (an der Rückenflosse) der Nerv der hinteren Extremität verlagert. An dieser Stelle, welche als Rückenregion nicht fähig ist eine Extremität zu bilden, wurde ein junges regenerierendes Blastem einer vorderen Extremität zusammen mit einem grossem Stück der es umgebenden Haut transplantiert. Das Blastem und die Haut wurden sorgfältig von den unter ihnen liegenden anderen alten Geweben des Organrestes abgetrennt. In verschiedenen Versuchsserien wurden unter das Blastem verschiedene Gewebe implantiert, welche als mögliche Determinations-Quellen untersucht wurden. Die Versuche wurden mit Axolotl (*Amblystoma mexicanum*) im Alter von 10 Monaten bis $1^1/_2$ Jahren ausgeführt.

I. In dieser Kontroll-Versuchsserie wurden nur regenerierende Blasteme der vorderen Extremität zusammen mit einem Stück der Haut transplantiert. Die Blasteme wurden in einem Alter von 10, 17, 20 und 22 Tagen genommen, jedoch so, dass sie dem Entwicklungsstadium nach eine kleine, nicht vorgewölbte Knospe darstellten.

Von 10 operierten Fällen blieben bis zum Versuchsende 9. In 2 Fällen kam es nicht zur Bildung irgendeines Organs. In einem Fall bildete sich ein kleines Organ von unbestimmtem Typus mit fingerartigem Auswuchs

und in den übrigen 6 Fällen kam es zur Bildung von flossenartigen Organen.

Aus diesen Ergebnissen können wir schliessen, dass die untersuchten regenerierenden Blasteme undeterminiert waren.

II. In der zweiten Versuchsserie wurden unter das transplantierte regenerierende Blastem der vorderen Extremität fein zerkleinerte, weiche mesodermale Gewebe der vorderen Extremität (Muskeln und Bindegewebe) implantiert. Wie aus den Arbeiten von Lusner und Woronzowa (?) bekannt ist, üben diese Gewebe einen wesentlichen Einfluss auf den Charakter der Regeneration eines Organs aus. Die vorliegende Versuchsserie hatte den Zweck, einen möglichen Einfluss der topographischen Struktur dieser Gewebe durch die Destruktion der Gewebe auszuschliessen.

Die Blastema wurden in dem gleichen Entwicklungsstadium wie in der ersten Versuchsserie im Alter von 8, 13, 14 und 22 Tagen genommen.

Von 10 Tieren der ersten Gruppe dieser Versuchsserie fielen 2 sehr bald weg. In einem Fall erfolgte eine völlige Resorption des Transplantats und das Transplantationsgebiet begann zu vernarben. In einem Fall erfolgte eine Resorption des Transplantats, und an seiner Stelle kam es zur Bildung eines flossenähnlichen Organs. In den 6 übrigen Fällen entwickelte sich aus der Transplantatanlage typische vordere Extremitäten. Die Entwicklung derselben war verschieden und lag wahrscheinlich von der Lage ab, welche der Extremitätenanlagen bei der Transplantation erhielt wurden.

In der zweiten Gruppe dieser Serie wurden 10 Operationen ausgeführt. In 8 Fällen fielen die Blastema der Transplantate ab. Nachdem wenige kam es an der Transplantationsstelle zu Organbildungen. In 5 Fällen wurden diese in frühen Entwicklungsstadien fixiert. In den 3 übrigen Fällen bildete sich: in einem Fall eine typische vordere Extremität, in zweiten Fall eine missgestaltete Extremität, welche mit flossenartigen Geweben vereinigt war, im dritten Fall ein Organ unbestimmten Typus mit 2 Fingern, im vierten Fall ein abgeflachtes flossenartiges Organ und im fünften Fall nichts gebildet war.

Diese Ergebnisse erlauben darauf zu schliessen, dass in Experimenten die Determination der regenerierenden Anlage unter Anwesenheit weicher mesodermaler Gewebe vor sich geht. Hierbei hat die topographische Struktur dieser Gewebe keine Bedeutung. Da sich in einigen Fällen die Extremitäten aus diesen weichen mesodermalen Geweben bildeten im Falle des Abfalls der über ihnen liegenden Blastema, so bedeutet das, dass sie ausser dem Epithel des Regenerats die Quelle des Regenerationsmaterials waren. Diese Ergebnisse erlauben jedoch noch nicht festzustellen, ob diese Gewebe noch nicht eine andere Rolle spielen.

III. Die dritte Versuchsserie hatte den Zweck, die Eigenschaften des Zellmaterials der determinierten regenerierenden Anlage festzustellen.

Die Blastema wurden in einem Alter von 11 und 13 Tagen transplantiert. Unter das Blastem wurden zerkleinerte, subepitheliale Teile von 3 bis 4 determinierten regenerierenden Blastemen im Alter von 30 Tagen, die sich im Stadium der Konusbildung befanden, implantiert.

In 2 von 4 operierten Fällen kam es zur völligen Resorption der Transplantate. Im dritten Falle bildete sich eine Extremität, welche mit flossenartigen Geweben vereinigt war. Im vierten Fall bildete sich ein grosses Organ unbestimmter Art, das anscheinend Extremitätenelemente enthielt.

IV. Die vierte Versuchsserie sollte klarstellen, ob das Material der undeterminierten regenerierenden Blastema zur Selbstorganisation fähig ist. In dieser Versuchsserie wurde der Fall betrachtet, dass das undeterminierte

bei den regenerierenden Blastemen seine Unfähigkeit zu Wachstum und Entwicklung infolge fehlenden Zuwachs determinierenden Zellmaterials zur Anlage.

Es wurden junge Blastneme im Alter von 11, 12 und 13 Tagen transplantiert. Unter diesen wurden verschiedene entsprechende Blastemstadien des gleichen Alters implantiert. Da man die Bewegung der Transplantate und ihre Verdrängung durch irritierte Gewebe in Betracht ziehen muss, wurden für zu implantierenden Gewebe von 9 oder 11 Blastemen genommen.

Diese Operationen erwiesen sich technisch als sehr schwierig. Von 19 Fällen verschwanden 17 Transplantate und fielen ab. In den 2 übrigen Fällen heilten die Blastneme gut ein; in beiden Fällen entstehen vom Extremitätchen. Ein Tier wurde 2½ Monate nach dem Versuch fixiert. Bei diesem hatte sich aus dem Transplantat eine typische vordere Extremität entwickelt. Bei dem anderen Tier wurde das Transplantat am 26. Tag nach dem Versuch fixiert. Das Transplantat stellte eine typische regenerierende Extremitätsanlage in Form einer kleinen Schaufel mit erkennbarer Zackenbildung der Finger dar. Die histologische Untersuchung ergab, dass im Innern der Anlage eine Anhäufung und Verknorpelung der Skeletelemente vor sich geht und dass die Anlage reich mit Blutgefässen versorgt ist. An der Basis der Anlage befindet sich eine kompakte Masse gleichartiger Zellen von embryonalem Typus der entsprechenden Teile des regenerierenden Blastems. Diese Zellen liegen sehr dicht aneinander.

Man kann annehmen, dass die Determination der Regenerate der Extremität in zwei Etappen vor sich geht. In der ersten Etappe kommt es zur Anhäufung von Regenerationsmaterial, in der zweiten unterliegt es irgendwelcher qualitativ verändernden Einwirkung. Man kann zum Beispiel annehmen, dass Bindegewebe als Material für die entsprechenden Teile der regenerierenden Anlage dient, während die Muskeln des Organismus einen bestimmten, spezifisch determinierenden Stoff absondern. Die fünfte Versuchsserie hatte den Zweck, diese Frage zu klären.

Bei der Transplantation wurden unter der regenerierenden Anlage weiche mesodermale Gewebe der vorderen Extremität (Muskeln und Bindegewebe) implantiert, welche durch Kochen oder Austrocknung abgetötet waren.

Im ganzen wurden 9 Operationen ausgeführt. In 4 Fällen wurden Blastneme im Alter von 10 Tagen und in 5 Fällen im Alter von 14 Tagen genommen. In diesen letzteren Fällen waren die Blastneme schon etwas weiter entwickelt als in den vorherigen Versuchsserien, und es war daher nicht ganz sicher, ob sie nicht schon determiniert waren.

In 3 Fällen entwickelte sich überhaupt nichts. In 3 anderen Fällen entwickelte sich ein Organ von undeterminiertem Typus, welches einige Elemente einer Extremität (Finger, Skeletteile) besass; und in einem Fall bildete sich eine gestaltete Extremität. Leider lassen diese Ergebnisse nicht die Möglichkeit bestimmte Schlussfolgerungen zu ziehen, da die Blastneme der 3 letzteren Fälle in einem Alter von 14 Tagen schon weiter entwickelt waren als in den übrigen Versuchen. Daher können diese Ergebnisse damit erklärt werden, dass bereits determinierte Blastneme transplantiert wurden.

Schlussfolgerungen

Es wäre verfrüht, auf Grund des hier angeführten Materials bestimmte Schlussfolgerungen zu machen, jedoch ist es immerhin möglich, einige Folgerungen zu machen, welche für die Anstellung weiterer Untersuchungen Bedeutung haben können.

1. Im Experiment nehmen weiche mesodermale Gewebe des Organrestes (Muskeln und Bindegewebe) unmittelbar Anteil bei der Determination des Regenerats der Extremität, hierbei hat die normale topographische Struktur dieser Gewebe keine Bedeutung. Diese Gewebe bilden die Quelle des Regenerationsmaterials.

2. Das Regenerat der Extremität kann determiniert werden, wenn man als Quelle des Regenerationsmaterials zerkleinerte Gewebe mesodermaler Teile von determinierten und auch nicht determinierten regenerierenden Extremitätenanlagen nimmt. Anscheinend ist dieses Material potentiell mit den weichen mesodermalen Geweben des Organrestes identisch.

3. Die Undeterminiertheit des jungen Regenerats der Extremität ist durch die Unterbrechung der Verbindung zwischen diesem und Geweben des Organrestes bedingt und liegt anscheinend vor allem in der Ausschliessung der Quelle des regenerierenden Materials. In diesem Fall muss die primäre Anhäufung von Zellmaterial in der regenerierenden Anlage hauptsächlich auf nicht mitotischem Wege vor sich gehen.

4. Die Vermutung ist nicht auszuschliessen, dass alte mesodermale Gewebe des Organrestes, welche die Quelle von regenerierenden Material bilden, noch irgendeine zusätzliche, vielleicht chemische und spezifische Einwirkung auf das sich bildende Regenerat ausüben. Ebenso ist es nicht ausgeschlossen, dass es unter besonderen Bedingungen möglich ist, die Entwicklungsrichtung des Zellmaterials der jungen, regenerierenden Anlage zu ändern.

Laboratorium für experimentelle Zoologie.
Akademie der Wissenschaften der UdSSR.
Moskau.

Eingegangen
d. 3. II. 1937.

ZITIERTE LITERATUR

[1] L. D. Liosner u. M. A. Woronzowa, Zool. Anz., **110** (1935); **115** (1936).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

EXPERIMENTAL ZOOLOGY

ON SOME DERIVATIVES OF THE EPITHELIAL FRAMEWORK OF THE THYMUS GLAND OF AN AMPHIBIAN

By A. M. VASJUTOČKIN

(Communicated by I. I. Schmalhausen, Member of the Academy, 29. III. 1937)

Some authors are inclined to believe that the inability of modern physiology to explain the functions of the thymus gland is due to the fact that histologists have not yet attained a clear understanding of the structure of that organ.

In fact histological investigations always stumble against the complexity and intricacy of the structure of that gland. It might seem that this question is exceedingly simple. As Hammar (1905) has shown, the framework of this organ consists primarily of a network built up of epithelial cells, and secondarily of lymphoid elements filling the interspaces and causing the differentiation of this organ into the cortical and medullary layers. But its structure is further complicated and obscured by the presence in the parenchyma of such elements which seem to be entirely «unnecessary», «superfluous» and consequently unintelligible. To these belong, in the first place, the myoid or muscle-like elements described by me in a series of papers (1913—1922). Of as great an interest are the ciliated, goblet-shaped and granular cells, the Hassall's corpuscles, and also the formations known under the name of «epitheliale Zellverbände» («placards épithéliaux» of the French authors), which can originate the so-called «cysts» of the thymus.

These cavities and ducts in the thymus gland, which are sometimes lines with ciliated cells, were long considered to be embryonic remnats. This hypothesis about an «abnormal development of the organ with the preservation of the primary duct» is very typical of works of a purely morphological character, which disregard the functional modifications of an organ. The fallacy of such a conception becomes apparent as soon as we put the question: Why does the quantity of these «embryonic remnants» in the thymus gland of an amphibian markedly increase in winter, while in summer they are relatively seldom met with in the frog?

The cysts of the thymus gland present a most varied picture. When examining serial sections, one easily establishes the fact that the cysts originate mostly in the layer, either blindly, as an aggregation of epithelial cells, or they terminate at the lacunae formed by the wide separation of the tissue elements of the thymus gland. In the most markedly developed cases, this cavity gradually extends to the cortical region, pierces it, and

even continues outside this organ into the connective-tissue capsule, where it again terminates blindly. In the majority of cases, however these cavities terminate within the medullary layer.

The structure of the walls of these cavities is likewise extremely varied. Sometimes the walls of these cavities are lined with a one-layered prismatic epithelium in whose cell protoplasm granules of secretion occur. There are cavities lined with a many-layered among whose elements one finds goblet cells secreting mucus; in places the wall is formed of a syncytial, or (after V. J. Rubashkin's terminology) symplastic mass.

Further, it should be noted that the free surface of the epithelial cells is frequently furnished with a cuticular rim. In other cases, one finds something like a ciliated fring resembling in structure that of the epithelium of renal ducts. As for the structure of the ciliated apparatus of the cells of the thymus gland, in some cases we have typical cilia with basal bodies; in other cases, the cilia are devoid of basal bodies or are furnished with numerous basal bodies; lastly there occur basal bodies alone. Therefore, complete agreement with what A. Gurvich (1901) stated in his description of the histogenesis of the ciliated epithelium is to be observed here. It is interesting to note that in the cells forming the wall of a large cavity there occur intracellular vacuols with intracellular cilia. As regards the contents of these cavities, it consists mainly of decomposition products of cell elements which appear to undergo autolysis by mucous secretion and leukocytes.

The question as to the significance of these formations naturally arises. The supposition that we have to deal here with a «survival» of the epithelial rudiment of the organ must, of course, be discarded. If we turn to the literature, an excellent summary of which was given by Hammar in 1909, we see that the development of these cavities is more or less associated with the processes of the involution of the organ, i. e. with certain degenerative processes. In fact, during the involution of the thymus gland, the difference between the cortical and medullary layers disappears, the organ acquires a «more epithelial» aspect, and a layer of prismatic cells arranged in the shape of a palisade appears on its surface.

In his «Biological Principles of Zoology», V. M. Shimkevich says that in the so-called «economic reduction», i. e. in the simplification of organs occurring under unfavourable conditions, we have to deal with a process similar to that of the «rejuvenation» of the cells which somewhat resemble embryonic elements, in other words, involution of the thymus gland would seem to furnish an example of the so-called «reversion to embryonic state». Such a conclusion, however, would be erroneous. It is to be supposed that differentiation of cells depresses, to a certain extent, their powers of multiplication, and that, when this differentiation is lost, the power of these cells to multiply again manifests itself. This lead the authors to the erroneous notion of the «rejuvenation of cells», or of their «reversion to embryonic state».

Of course, there is no doubt that, during the involution of the thymus gland, there occurs a partial de-differentiation of the epithelial cells of this organ. Owing to their cytotactic properties, these cells, on drawing together, form what we call epithelial cell aggregations. Under the influence of the decomposition products formed during the involution and acting analogously to necrohormones, these aggregations receive an impulse to progressive development. They strive, so to speak, to physiologically regenerate the tissue.

We know that tissue regeneration occurs after the type of embryonic

growth. In this connection it suffices to recall the regeneration of muscular tissue of, say, the *Caudata*, when typical myoblasts develop from disintegrating muscle fibre and originate new muscle fibres. Thus, it is not surprising that, during the formation of epithelial cavities in the thymus gland, we observe a reproduction of the structure the organ had at the time of its formation in the embryonic period.

However, V. M. Shimkevich (1923) notes a certain, as yet insufficiently investigated phenomenon, namely, the regeneration of an organ in the form that this organ had in the ancestors of this species (the so-called «atavistic phenomena in regeneration») and not in the form in which it existed in the individual in question. If we accept this standpoint, the presence in the thymus gland of ciliated and goblet cells, as well as of cells with a ciliated fringe, etc., will appear to be, instead of a histological curiosity, a «histological atavism». Thus, contrary to old authors, it is to be believed that the process of the formation of cavities in the thymus gland is not a degenerative, but a regenerative one. It is undoubtedly associated with the process of degeneration of the cell elements, but is itself of a opposite nature, as is proved by the presence of karyokinetic cell division in the walls of these cavities. Furthermore, it is difficult to believe that a cell of the epithelial framework should, in degenerating, change into a ciliated or secretory cell. The development of these structures can be explained only by the de-differentiation of the epithelial cells and by their subsequent complication under the influence of the partly disintegrating tissue of the thymus gland.

Histological Laboratory.
Third Leningrad Medical Institute.
Leningrad.

Received
29. III. 1937.

REFERENCES

[1] J. Aug. Hammar, Anatomischer Anz., 27 (1905). [2] J. Aug. Hammar, Fünfzigjahre Thymusforschung. Ergebnisse der Anatomie und Entwicklungsgeschichte, Bd. XIX (1909). [3] A. P. Dustin, Archives de Biologie, XXVIII (1913). [4] В.М. Шимкевич, Биологические основы зоологии, Ленинград (1923).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

EXPERIMENTELLE ZOOLOGIE

DIE MORPHOGENETISCHE AKTIVITÄT DER VERSCHIEDENEN TEILE DER HYPOPHYSE

III. DER EINFLUSS DER VERSCHIEDENEN ZONEN DES HYPOPHYSEVORDERLAPPENS AUF DIE SCHILDDRÜSE DER AMPHIBIEN

Von A. A. VOITKEVIČ (WOITKEWITSCH)

(Vorgelegt von I. I. Schmalhausen, Mitglied der Akademie, d. 14. IV. 1937)

In den vorhergehenden Mitteilungen wurde der morphogenetische Effekt bei Implantation verschiedener Teile des Hypophysevorderlappens, welche sich ihrer histologischen Struktur nach unterscheiden, beschrieben. Die Überpflanzung von Gebieten der «Basophilen Zone» an Kaulquappen und Axolotl ruft beschleunigte Metamorphose hervor, während die «Eosinophile Zone» das Wachstum stimuliert. Diese Befunde stehen im Widerspruch mit den Angaben von Spaul (1930), welcher bei analogen Versuchen mit verschiedenen Teilen des Hypophysevorderlappens metamorphogene Wirkungen erhielt, die sich nur durch die Intensität unterschieden.

Wir haben schon früher die Ansicht geäussert, dass die Beschleunigung der Metamorphose mit Substanzen aus der «Basophilen Zone» mit dem in diesem Teil der Hypophyse lokalisierten thyreotropen Hormon im Zusammenhange steht, welche die Schilddrüse der Versuchstiere (Kaulquappen) aktiviert. Spaul (1930) vertritt die Meinung, dass die Metamorphose der Axolotl durch die Einführung von Substanzen des Hypophysevorderlappens hervorgerufen werden kann, unabhängig von der Schilddrüse. Jedoch in einer Reihe von Arbeiten anderer Forscher, und besonders Uhlenhuths, wird eine andere Anschauung vertreten, u. zw. dass die Metamorphose der Amphibienlarven bei Einführung der Vorderlappen gefördert wird, indem die Schilddrüse aktiviert wird. Da der Einfluss der einzelnen Teile des Hypophysevorderlappens auf die Schilddrüse der Amphibienlarven bis jetzt noch nicht genau erforscht wurde, so haben wir die histologische Untersuchung des Materials (Kaulquappen von *Rana temporaria*) der in der ersten Mitteilung beschriebenen Versuche durchgeführt. Wir wiederholen kurz die Resultate diser Versuche. Klarer Effekt, d. h. Beschleunigung der Metamorphose, wird schon 5 Tage nach der Implantierung der «Basophilen Zone» beobachtet: scharfe Verringerung des Gesamtgewichtes der Kaulquappen, der Darmkanal erfährt eine Verkürzung bis zum Minimum, der Schwanz wird bis zur Hälfte resorbiert (nach Gewicht), die vorderen Extremitäten brechen durch und die Entwicklung der hinteren Extremitäten wird stimu-

liert. Bei der Implantierung der «Eosinophilen Zone» beobachten wir Vergrösserung der Gesamtgrösse der Kaulquappen und aller untersuchten Organe der Larven. Zu genanntem Zeitpunkt, d. h. 5 Tage nach der Implantierung, wurde von den Kaulquappen aller Serien das Gebiet des Halses, welcher die Schilddrüse einschliesst, fixiert und Serialschnitte von 7 Mikron hergestellt (Färbung Hämalaun-Eosin und Asakarmin-Mallory).

Auf dem grössten Schnitt jedes Teils der Schilddrüse wurde die Anzahl der Follikeln gezählt, die Epithelhöhe und der innere Durchmesser aller Follikeln gemessen; auch die Grösse und Fläche des ganzen Schnittes wurde festgestellt. Die so erhaltenen Resultate wurden summiert und das Mittel für jede Serie bestimmt (Tabelle).

Die Charakteristik des Maximalschnittes durch ein Kaulquappenschilddrüse

Serie	Länge der Drüse in Mikronen	Breite der Drüse in Mikronen	Fläche in qu. Mikronen	Zahl der Follikeln	Epithelhöhe in Mikronen	Innerer Durchmesser d. Follikeln in Mikronen
Kontrolle	354	139	190 793	18.5	4.12	27.80
I. Eosinophiler Teil .	309	150	165 387	24.0	5.68	18.42
II. Basophiler Teil . .	407	286	376 994	20.5	10.32	15.46

Aus der Tabelle ist zu ersehen, dass die Grösse der Schilddrüse* der Kaulquappen, welchen die «Eosinophile Zone» implantiert wurde (I. Serie),

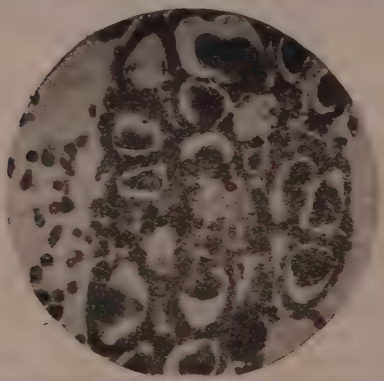


Abb. 1. Die Schilddrüse der Kontrollkaulquappen. Objektiv 4 mm, Okular 7K.

Abb. 2. Schilddrüse der Kaulquappen der I. Serie (Implantation der «Eosinophilen Zone»). Objektiv 4 mm, Okular 7K.

sich praktisch von der Kontrolle nicht unterscheidet. Bei der Implantierung von Geweben der «Basophilen Zone» (II. Serie) sind die Ausmasse des Organs bedeutend vergrössert—ungefähr zweifach.

* Genauer, die Grösse des maximalen Schnittes (s. oben).

Im histologischen Bau der Schilddrüse bemerken wir ebenfalls scharfe Unterschiede (Abb. 1—3). Die Epithelhöhe der Follikeln erreicht bei Implantierung der «Basophilen Zone» im mittleren 10.32 Mikron, was sogar für die Periode der höchsten Aktivität dieses Organs während des natürlichen Umwandlungsprozesses aussergewöhnlich ist. Die Ausmasse der Epithelzellen bei der Überpflanzung der «Eosinophilen Zone» sind bedeutend kleiner (5.68 Mikron) und unterscheiden sich wenig von der Kontrolle (4.12 Mikron). Der Durchmesser der Follikeln ist jedoch bedeutend kleiner und ihre Grössen nähern sich denen der II. Serie. Somit sehen wir, dass der Thyreoideapparat der Versuchskaulquappen der I. Serie, so wie auch bei der Kontrolle, sich in einem relativ inaktiven Zustand befindet, die Färbungsfähigkeit des Kolloids ist in beiden Fällen gleich. Der Zustand der Schilddrüse der Larven

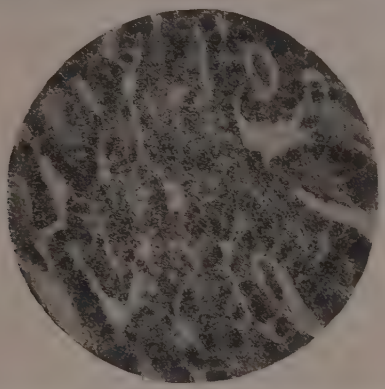


Abb. 3. Schilddrüse der Kaulquappen der II. Serie (Implantation der «Basophilen Zone»). Objektiv 4 mm, Okular 7K.

der II. Serie (Implantation der «Basophilen Zone») dagegen charakterisiert sich durch stark ausgeprägte Hyperfunktion. Ausser der Vergrösserung der Ausmasse der Drüse und der Epithelhöhe, was sich mit der Verringerung der Follikelgrössen vereinbart, ist die Konfiguration der Letzteren stark verändert; sie liegen parallel der Längsachse des Organs. Das Kolloid ist fast vollständig resorbiert, die stark vakuolisierten Reste färben sich blau und hellblau (Asan-Mallory). Alle Merkmale einer intensiven Funktion sind vorhanden.

Die erhaltenen Angaben zeigen vor allem, dass die basophilen Elemente die Träger des thyreotropen Hormons sind, während die Substanz der «Eosinophilen Zone» solche Eigenschaften nicht aufweist. Über diese Frage bestehen in der Literatur bis jetzt noch Meinungsverschiedenheiten [wir meinen hier nicht die früheren Arbeiten von Smith (1925), welche von Uhlenhuth kritisch analysiert wurden]. Auf Grund der Erforschung des Baues der Hypophyse während der Metamorphose kommt Klemens (1932) zu dem Schluss, dass die Beschleunigung derselben von den eosinophilen Elementen abhängt. Die nach anderer Methode erhaltenen Resultate von Lebedeva (1936) sprechen jedoch gegen die Lokalisierung des thyreotropen Hormons in den eosinophilen Zellen, da nach der Thyreoidektomie diese Elemente verschwinden. Bei Testierung solcher Hypophysen an Meerschweinchen zeigte Zekwer (1936) eine Verstärkung der thyreotropen Funktion (in bezug auf die ganze Hypophyse), was

auch sehr wahrscheinlich macht, dass die thyreotrope Funktion mit den basophilen Zellen in Zusammenhang steht.

In unseren Versuchen wurde die Testierung der «Basophilen Zone» einzel ausgeführt, welche die Richtigkeit der genannten Voraussetzung direkt bestätigte.

Hierdurch wird auch sehr wahrscheinlich, dass die bei der Implantierung der «Basophilen Zone» beobachtete Metamorphose nur die Folge der Aktivierung der eigenen Schilddrüse der Larven ist, d. h. die Wirkung der Hypophyse ist hier nur indirekt. Die Versuche von Schwarzbach und Uhlenhuth (1933), sowie auch Figge und Uhlenhuth (1933) zeigten, dass es unmöglich ist einen morphogenen Effekt vom Hypophysevorderlappen an thyreoidektomierten Larven zu erhalten. Die Wiederholung dieses Versuches an unserem Material und nach unserer Methodik gibt die Möglichkeit die hier gestellte Frage endgültig zu lösen.

Laboratorium für Entwicklungsmechanik.
Akademie der Wissenschaften der UdSSR.
Moskau.

Eingegangen
d. 14. IV. 1937.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

EXPERIMENTELLE ZOOLOGIE

DIE MORPHOGENETISCHE AKTIVITÄT DER VERSCHIEDENEN TEILE DER HYPOPHYSE

IV. ÜBER DIE HEMMUNG DER METAMORPHOSE DER KAULQUAPPEN DURCH SUBSTANZ DER «EOSINOPHILEN ZONE» DES HYPOPHYSEVORDERLAPPENS

Von A. A. VOITKEVIČ (WOITKEWITSCH)

(Vorgelegt von I. I. Schmalhausen, Mitglied der Akademie, d. 14. IV. 1937)

Ungeachtet dessen, dass schon eine grosse Anzahl von Arbeiten über den Einfluss des Hypophysevorderlappens auf das Waschstum und die Entwicklung der Amphibien vorliegt, bestehen in manchen Fragen noch Widersprüche, die einer weiteren Klärung bedürfen. Dies bezieht sich vor allem auf die Versuche über die Einführung von Substanzen des Hypophysevorderlappens an Larven.

Einige Autoren beobachteten Wachstumsstimulierung, die anderen dagegen—Beschleunigung der Metamorphose. Auf eine der möglichen Ursachen dieser Widersprüche hat Spaul (1924—1930) hingewiesen, welcher Beschleunigung der Metamorphose unter dem Einfluss eines Präparates des Vorderlappens (Fa. Armour u. Co), welches grossen Jodgehalt aufwies. erhielt. Weiter hat Spaul (1930) festgestellt, dass die Extrakte aus einem ganzen Hypophysevorderlappen, welche im Laboratorium mit schwacher Essigsäurelösung hergestellt wurden, bedeutende morphogene Aktivität besitzen, die sich aber mit Veränderung der Säurekonzentration verändern. Extrakte, die mit Ringerlösung, Azeton u. a. hergestellt wurden, sind inaktiv. Die Autolyse der Hypophysegewebe führt zur Verminderung der metamorphogenen Aktivität, hängt aber auch wieder ab von der Temperatur, bei der die Autolyse vor sich geht.

Die angeführten Angaben weisen auf die Wichtigkeit der Herstellungsweise der Hypophysepräparate hin (in bezug auf ihre biologische Aktivität) und die Art der Aufbewahrung des Rohstoffes vor der Bearbeitung und können zum Teil zur Klärung der Widersprüche in den Resultaten der einzelnen Autoren dienen.

Ausserdem ist auch sehr wahrscheinlich, dass die Unterschiede im Effekt mit den Veränderungen des physiologischen Zustandes der testierten Hypophyse im Zusammenhang stehen, was sich an deren Bau äussert. Es ist bekannt, dass das Verhältnis der zwei Hauptelemente des chromophilen Teils des Hypophysevorderlappens—Eosinophile und Basophile—in Abhängigkeit von einer ganzen Reihe von Umständen sich scharf ändern (Kahn,

1936). Bei der Herstellung der Extrakte aus einer ganzen Drüse kommt ein neuer variierender Faktor hinzu, welcher den Unterschied in der biologischen Aktivität des Präparates bestimmen kann. Grosse Bedeutung haben da die Untersuchungen von Smith u. Smith (1923), welche die basophilen und eosinophilen Teile der Hypophyse bei der Testierung an Kaulquappen isolierten. Im ersten Falle erhielten sie eine Beschleunigung der Metamorphose und im zweiten—Wachstumsstimulierung.

Spaul (1930) erforschte ebenfalls einzeln die morphogenetischen Eigenschaften der verschiedenen topograhischen Gebiete des Vorderlappens, indem er ihn in drei Gebiete teilte: äussere, mittlere und innere. Eine solche Teilung kann nicht als gelungen angesehen werden, da sie nicht ganz den Besonderheiten der topographischen Lage der basophilen und oxyphilen Elemente entspricht. Zum Schluss kommt der Autor zu einer Anschauung, die der von Smith widerspricht und bringt die morphogene Wirkung mit den oxyphilen Elementen in Zusammenhang. Wie schon in unserer ersten Mitteilung gezeigt wurde, ist es richtiger in der Hypophyse des Stieres nur zwei Zonen zu unterscheiden. Der Zentralteil, welcher ein kleines Gebiet einnimmt, besteht praktisch nur aus Basophilen. Er sendet aber viele Abzweigungen nach der Peripherie, wo hauptsächlich nur Eosinophile vorhanden sind. Wenn wir zu den Versuchen nicht die ganze Drüse, sondern nur die weit von einander entfernten Gebiete der Hypophyse verwenden, wird es möglich, die morphogene Wirkung der «Basophilen» und «Eosinophilen Zone» isoliert zu erforschen. Auf Grund der Implantationsversuche an Kaulquappen (unsere erste Mitteilung) kommen wir in Übereinstimmung mit Smith (1923) zu dem Schluss, dass die Fähigkeit, die Metamorphose zu aktivieren nur die «Basophile Zone» besitzt, die «Eosinophilen» stimulieren das Wachstum, wodurch auch der Umwandlungsprozess gehemmt wird. Der letztere Effekt kann in klarerer Form natürlich bei beschleunigter Metamorphose erhalten werden. Ausgehend von solchen Voraussetzungen haben wir die Wirkung der Implantation der «Eosinophilen Zone» bei Aktivierung der Metamorphose mit Schilddrüse untersucht. Unter diesen Verhältnissen muss sich die Hemmung der metamorphotischen Prozesse klar äussern. In der vorliegenden Arbeit haben wir folgende Versuche mit Implantation von Substanzen der «Eosinophilen Zone» ausgeführt:

1. Implantation an Kaulquappen, welche im Kulminationsstadium der Metamorphose waren (und wo die eigene Schilddrüse stark funktionierte).

2. Die Implantation wurde an Kaulquappen ausgeführt, deren Metamorphose in einem frühen Stadium experimentell durch Aktivierung der eigenen Schilddrüse beschleunigt wurde, durch Überpflanzung von Substanz aus der «Basophilen Zone» (s. III. Mitteilung dieser Serie).

3. Die Metamorphose der Kaulquappen wurde beschleunigt, indem diese in eine Emulsion von Schilddrüsenpräparat gebracht wurden.

Die Stückchen der «Eosinophilen Zone» (Gewicht 1.5—2.0 mg) wurden wie auch früher den Kaulquappen *Rana temporaria* in die Bauchhöhle implantiert. Für den Versuch wurden die an der äussersten Peripherie liegenden Teile des Hypophysevorderlappens ausgenützt, in welchen die Zahl der Basophilen Elemente am niedrigsten ist, was auch parallel durch mikroskopische Untersuchung bestätigt wurde. Die Teile der «Basophilen Zone», die in der zweiten Gruppe verwendet wurden, entnahm man dem Zentrum, wodurch das Vorhandensein von eosinophilen Zellen ausgeschlossen wurde.

Als Kontrolle dienten normale Kaulquappen desselben Alters. Jede Gruppe bestand aus 15—25 Individuen. Sobald zwischen Versuchs- und Kontrollarven ein klarer Unterschied festgestellt werden konnte, wurden

alle gleichzeitig getötet. Wie auch früher wurde das Gesamtgewicht, das Gewicht des Schwanzes und der hinteren Extremitäten, wie auch die Länge des Gesamtkörpers, Schwanzes und Darmkanals berechnet.

In den hier gebrachten Tabellen werden die mittleren Zahlen angeführt. Der Kürze halber bringen wir nur einen Teil der Angaben, und lassen die Doppelserien aus.

In Tab. 1 sind die Angaben des Versuchs an natürlich metamorphosierenden Kaulquappen angeführt.

Tabelle 1

Der Einfluss der Substanz der «Eosinophilen Zone» auf die Entwicklung der Kaulquappen, welche sich im Kulminationsstadium der Metamorphose befanden

Tage nach Versuchs Beginn	Serie	Gewicht in mg			Länge in mm		
		Ganze Kaulquappe	Schwanz	Hintere Extremität	Ganze Kaulquappe	Schwanz	Darmkanal
0	Ausgangsmaterial	701.0	145.0	7.3	41.9	27.0	166.3
10	Kontrolle . . .	321.3	57.8	21.4	29.2	17.8	45.3
	«Eosinophile Zone»	625.5	134.0	17.8	41.1	28.2	115.2
15	Kontrolle . . .	292.7	5.2	29.1	14.2	2.2	26.0
	«Eosinophile Zone»	415.5	70.4	19.6	31.0	18.0	49.0

Die Vergleichung der Kontroll-Kaulquappen mit dem Ausgangsmaterial zeigt schon zum 10. Tage Merkmale einer Metamorphose, welche zum 15. Tage endet (s. Masse des Darmkanals und Schwanzes). Bei den Versuchskaulquappen ist die Metamorphose stark gehemmt, aber doch nicht vollständig.

In den Versuchen der zweiten Gruppe wird gezeigt, dass die Substanz aus der «Eosinophilen Zone» die Metamorphose auch dann hemmt, wenn diese experimentell durch die Aktivierung der eigenen Schilddrüse hervorgerufen wird (Tab. 2).

Die Beschleunigung der Metamorphose wird, wie auch früher bei der Überpflanzung der «Basophilen Zone» allein, sehr klar demonstriert: das Gesamtgewicht der Larven fällt, und der Schwanz wird resorbiert. Der Effekt wird jedoch bedeutend vermindert bei gleichzeitiger Implantation von Stückchen der «Eosinophilen Zone». Es wird eine Art von «Neutralisierung» der morphogenetischen Wirkung der «Basophilen Zone» beobachtet. Wenn wir das Grössenverhältnis zwischen den Baso- und Oxyphilen «Zonen» im ganzen Vorderlappen einer Stierhypophyse einschätzen (die mittleren Zahlen, welche wir bei Gewichtsfeststellungen einiger Hypophysen erhielten, sind: «Basophile Zone» 86 mg, «Eosinophile Zone» 2 671 mg), so sehen wir, dass bei der Verwendung eines Extraktes aus der ganzen Drüse die morphogenetische Wirkung der «Basophilen Zone» sich entweder nur schwachem Grade äussern kann oder auch ganz ausbleibt.

Tabelle 2

Der Einfluss auf die Entwicklung der Kaulquappen bei gleichzeitiger Implantation von Substanzen der «Eosinophilen» und «Basophilen Zone» des Hypophysenvorderlappens (Angaben zum 7. Tag nach der Implantation)

Serie	Gewicht in mg			Länge in mm		
	Ganze Kaulquappe	Schwanz	Hintere Extremität	Ganze Kaulquappe	Schwanz	Darmkanal
Kontrolle	233.3	34.4	1.9	29.9	16.4	74.3
«Basophile Zone»	160.0	22.0	2.7	22.5	12.7	27.4
«Eosinophile» und «Basophile Zone»	187.7	24.6	2.1	22.8	13.7	39.8

In der dritten Gruppe wurde die Beschleunigung der Metamorphose durch die Unterbringung der Larven in eine Thyreoidinemulsion schwacher Konzentrierung hervorgerufen (Tab. 3).

Tabelle 3

Der Einfluss der Substanz aus der «Eosinophilen Zone» auf thyreoidisierte Kaulquappen

Konzentration des Thyreoidins	Serie	Gewicht in mg			Länge in mm		
		Ganze Kaulquappe	Schwanz	Hintere Extremität	Ganze Kaulquappe	Schwanz	Darmkanal
	Kontrolle . .	133.4	40.8	0.8	24.8	16.4	25.2
1:300 000	«Eosinophile Zone» . . .	166.4	39.4	1.3	26.3	17.9	30.8
	Kontrolle . .	190.0	37.5	1.1	29.0	17.2	68.7
1:500 000	«Eosinophile Zone» . . .	290.2	55.0	2.1	31.0	19.7	82.5
	Kontrolle . .	197.5	41.7	1.9	28.2	17.7	64,2
1:1000 000	«Eosinophile Zone» . . .	265.0	57.0	2.6	30.4	20.2	76.0

Die Vergleichung mit dem Ausgangsmaterial (die Angaben werden hier nicht angeführt) zeigt, dass die Kontrollkaulquappen, welche in Thyreoidin gehalten wurden, bedeutend metamorphosierten (besonders in der stärk-

sten Konzentration). Bei der Implantation von Stückchen der «Eosinophilen Zone» wird, wie auch früher, die Metamorphose bedeutend gehemmt, wobei sich der stärkste Effekt bei der schwächsten Thyreoidin-Konzentration äussert. In dieser Hinsicht stimmen unsere Resultate mit denen von Smith (1926) überein, welcher in Versuchen an Colorado-Axolotln ebenfalls Neutralisierung der metamorphogenen Wirkung beobachtete bei gleichzeitiger Einwirkung mit Schilddrüse und Extrakt aus der frischen Hypophyse.

Somit sehen wir, dass alle 3 Versuchsgruppen übereinstimmende Resultate zeigen, d. h. die Hemmung der Metamorphose bei Implantation von Substanzen aus der «Eosinophilen Zone» des Vorderlappens.

Dasselbe beobachteten wir auch in den von uns durchgeführten Ergänzungsversuchen, wo die Beschleunigung der Metamorphose durch äussere Faktoren hervorgerufen wurde. Wie schon Barfurth (1887) zeigte und Križenecky (1914) bestätigte, wird die Metamorphose durch Hunger, besonders in den späten Stadien sehr beschleunigt.

In Tab. 4 bringen wir die Resultate von 2 Versuchen, die sich durch die Dauer unterscheiden (7 und 12 Tage). Die Kaulquappen erhielten während dieser Zeit keine Nahrung, den Versuchstieren wurden Stückchen der «Eosinophilen Zone» transplantiert. In dieser Tabelle werden auch die Angaben eines dritten Versuches angeführt, wo die Beschleunigung der Metamorphose durch hohe Temperatur (25—26°) erreicht wurde. Siehe hierüber die Angaben von Hertwig (1894), Lilli (1897), Adler (1916), Belkin (1934), Woitkewitsch (1935) u. a.

Tabelle 4

Der Einfluss der Gewebe der «Eosinophilen Zone» des Hypophysevorderlappens auf die Entwicklung der Kaulquappen bei vollständigem Hungern und hoher Temperatur

Bedingungen	Serie	Gewicht in mg			Länge in mm		
		Ganze Kaulquappe	Schwanz	Hintere Extremität	Ganze Kaulquappe	Schwanz	Darmkanal
Hungerregime	Kontrolle . .	262.5	68.3	4.3	32.6	21.6	56.0
	«Eosinophile Zone» . . .	372.5	80.1	4.0	34.2	23.7	93.8
Hungerregime	Kontrolle . .	168.7	31.2	4.9	24.1	15.2	22.3
	«Eosinophile Zone» . . .	313.6	69.5	3.1	34.3	24.2	73.6
Temperatur 25—26°	Kontrolle . . .	158.1	40.6	1.4	26.1	16.6	70.4
	«Eosinophile Zone» . . .	264.0	53.6	3.0	31.2	20.4	77.0

Bei den Kontrollkaulquappen aller drei Gruppen verläuft die Metamorphose intensiv. Bei Implantation von Substanzen der «Eosinophilen Zone» wird sie gehemmt.

Auf diese Weise ist es möglich bei verschiedenen Variationen des Hauptversuchs die Eigenschaften der Substanz der «Eosinophilen Zone» des Vorderlappens—den Metamorphoseprozess zu hemmen—klar zu beweisen.

Dieser Effekt steht ohne Zweifel im Zusammenhang mit der Stimulierung des Wachstums, was ein Antagonismus der Differenzierungserscheinung ist, aus welcher sich der Umwandlungsprozess zusammensetzt. Somit bestätigen die hier beschriebenen Versuche die früher festgestellte Tatsache über die verschiedene Bedeutung der einzelnen Teile der Hypophyse im Entwicklungsprozess der Amphibienlarven.

Laboratorium für Entwicklungsmechanik.
Akademie der Wissenschaften der UdSSR.
Moskau.

Eingegangen
d. 14. IV. 1937.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1937. Volume XV, № 6—7

PALAEONTOLOGY

GIGANTELLA SAR. AND ITS STRATIGRAPHIC IMPORTANCE IN THE LOWER CARBONIFEROUS DEPOSITS OF THE MIDDLE COURSE OF ISHIM RIVER

By S. M. ANDRONOV

(Communicated by A. D. Archangelskij, Member of the Academy, 23. III. 1937)

The earliest evidence of the occurrence of the genus *Gigantella* Sar. in Kazakhstan is to be found in Krasnopolskij's paper (4). Later on, upon examination of his collection by other investigators, his identification was found to be erroneous, and this gave rise to the opinion that *Productus giganteus* Mart. was altogether absent from the lower carboniferous deposits of Kazakhstan. This view was subsequently refuted by a number of geologists working in Kazakhstan. Mention may be made, e. g., of Bespalov's paper (3) containing a reference to a find of *Gigantella* Sar. in the Visé deposits of Dzheskazgano-Atbassar region. In 1937, S. M. Andronov (1), mentions the occurrence of *Pr. (Gigantella)*+aff. *donaicus* Leb. var lata var. n. in the middle course of Ishim river.

Thus, although some evidence as to the presence of *Gigantella* Sar. in Kazakhstan has appeared in literature in recent years, none of the writers seem to have had at their disposal collections large enough to warrant a specific determination, and as, on the other hand, the stratigraphy of the coal measures of Kazakhstan has been but indifferently studied, the stratigraphic importance of *Gigantella* has so far remained obscure.

As a result of two years' work conducted by the writer, on behalf of «Glavugol», in the middle course of Ishim river in 1935—1936, the writer succeeded in getting together a smal collection of *Gigantella*, a record being kept of the particular stratum where a specimen was found. In addition to this, the collection also contains several specimens of *Gigantella* found in the district around Kiptchak-Sor by D. N. Burtzev, Chief Engineer of the Ishim office.

The collection contains in all 26 specimens of *Gigantella*, amongst which the writer has succeded in identifying 6 species*: *Pr. (Gigantella) donaicus* Leb., *Pr. (Gigantella) bisati* Paeck., *Pr. (Gigantella)* cf. *sarytcheffi* Paeck., *Pr. (Gigantella) ishimicus* n. sp., *Pr. (Gigantella) kiptchakensis* n. sp., *Pr. (Gigantella) tulensis* Bolkh. var. *kasachstanensis* var. nov.

* A description of which is being prepared for the press by the writer.

I take this opportunity of conveying my gratitude to M. A. Bolkhovitinova who has kindly revised my identifications.

Four species of *Gigantella* had been found in the districts around Dalni village and Mikhailovka village, while *Pr. (Gigantella) kiptchakensis* n. sp. and *Pr. (Gigantella) tulensis* Bolkh. var. *kasachstanensis* var. nov. come from the Kiptchak-Sor district. I am at a loss to determine the stratigraphic position of the latter, as the specimens were handed over to me without any detailed section and as I have never been at Kiptchak myself. I shall therefore deal only with those *Gigantella* that I found in the district round Dalni and Mikhailovka villages.

The following is a rough outline of the section of the coal measures where those specimens were found. Resting upon the crimson-coloured conglomerates and sandstones of the Upper Devonian is a calcareo-schistose carboniferous series. It is characterised by the following forms*: *Spirifer tornacensis* Kon., *Productus burlingtonensis* Hall., *Productus vaughani* M. Wood., *Productus magnus* White. On the basis of this fauna, I have classified this series as Tournai. Its thickness is 75 m.

Overlying this series is a series of calcareous sandstones, at the base of which lies a horizon of glauconitic sandstones containing *Spirifer plenus* Hall. and belonging to the uppermost strata of Tournai or Tournai-Visé horizons. Resting on these are grey or white porous limestones, in which (2—3 m above the sandstones containing *Sp. plenus* Hall.) I found near Mikhailovka village in 1936 three specimens of *Gigantella*, which were identified as *Pr. (Gigantella) primordialis* n. sp. In point of age, these porous limestones belong to the lower Visé horizon.

Overlying these is a series of fossiliferous sandstones and limestones containing amongst others *Spirifer grimesi* Hall. (very numerous), *Spirefirina laminosa* McCoy, *Spirifer* aff. *princeps* McCoy, *Productus lobatus* Sow. (very numerous), *Pr. Lobatus* var. *laqueatus* M. Wood., *Pr.* aff. *gröberi* Krenk. The whole of this fauna suggests that the series is of Visé age (*Seminula* zone, S_1 subzone of the English division). Resting on this series are finely laminated grey limestones containing *Pr. productus* Mart., *Pr. productus* var. *hispidus* M. Wood., *Productus undatus* Defr. It was here that *Pr. (Gigantella) ishimicus* n. sp. was discovered near Mikhailovka village in 1936. Higher up are green calcareous limestones containing an abundance of *Chonetes* as well as the fauna enumerated above. It was here that a second specimen of *Pr. (Gigantella) ishimicus* n. sp. was discovered near Mikhailovka village in 1936. In the same year, two specimens of *Pr. (Gigantella) bisati* Paeck. were found in these sandstones near Dalni village. The series just described corresponds in all probability to subzone S_2. The total thickness of the calcareous sandstone series is 200 m.

Resting on it is a thin horizon of reddish sandstones, which I take to be the boundary line between the Lower and the Upper Visé (between *Seminula* and *Dibunophyllum* zones). They are overlaid by a series of argillitic sandstones, consisting at its base of green pyritized sandstones and limestones, and, higher up, of grey crinoid limestones containing *Pr. gröberi* Krenk. and *Pr. (Gigantella) donaicus* Leb. var. lata nov. var. This *Gigantella* was found by the writer near Dalni village (Yuzhnaya creek) in 1935. In 1936, several defective specimens of the same species were discovered in the same locality.

Resting on the horizon of crinoid limestones with *Pr. (Gigantella) donaicus* Leb. var. lata nov. var. is a thousand-metre thick series of green of tobac-

* For a complete record of fauna cf. Andronov's article [1].

co-coloured argillites, alleurites and sandstones with occasional streaks of limestones containing fossils of *Dibunophyllum* (D) zone. The most careful search failed to reveal any *Gigantella* in this series.

Here ends the section of carboniferous deposits in the region under consideration.

If Paeckelman's classification of *Gigantella* were to be adopted, all the species found by the writer would be classed in the *Gigantoides* group, characterized by a very thin shell and the absence of secondary longitudinal plication. The *Gigantella* identified by the writer as *Pr. (Gigantella) bisati* Paeck. is strongly reminiscent of *Linoproductus* as regards general habitus and fine ribbing; an examination of its internal structure, however, induced the writer to place it amongst the *Gigantella*.

An examination of M. A. Bolkhovitinova's collection of Moscow-region *Gigantella** composed of specimens found by the owner in Tula seams, has revealed a striking resemblance between the Tula forms and certain specimens of our collection. Thus, *Pr. (Gigantella) bisati* Paeck. shows a great similarity to the form called *Gigantella primitiva* by M. A. Bolkhovitinova. Moreover, the forms found by D. N. Burtzev near Kiptchak-Sor and named by the writer *Pr. (Gigantella) kiptchakensis* n. sp. and *Pr. (Gigantella) tulensis* Bolkh. var. *kasachstanensis* var. nov. respectively are closely related to *Gigantella tulensis* Bolkh. ([2]) and are, to all appearances, varieties of that species.

As to the resemblance of Kasakhstan forms to the *Gigantella* of Serpukhov district, a detailed comparison of these forms with the originals described by T. G. Saryčova ([7]) has induced the writer to conclude that there is a great dissimilarity in the habitus of these forms. Unlike the *Gigantella* found by us, the specimens identified by Saryčova have an unmistakable longitudinal plication, a thicker shell and fully developed brachial cones.

A perusal of Chinese literature ([8]) showed that our forms had very little in common with the *Gigantella* of China.

There would thus seem to be a certain similarity between the Kasakhstan forms and the *Gigantella* of Germany and those of Tula seams of Moscow region basin.

As is known, *Gigantella* are mainly characteristic of *Dibunophyllum* (D) zone, but some writers mention the presence of primitive *Gigantella* in the upper portion of *Seminula* (S) zone. Thus, Vaughan ([10]) mentions the occurrence of *Productus* θ now referred to *Gigantella* in the *Seminula* (S) zone, *Pr. (Gigantella) donaicus* Leb. was encountered by Rotai ([6]) at the very top of C^2_1 5 series, which, according to Lisicin ([5]), corresponds to the *Seminula* (S) zone of the Anglo-Belgian basin.

As to the stratigraphic position of the *Gigantella* found in the middle course of Ishim river, it is to be noted that the genus *Gigantella* Sar. is encountered here at far greater depths than observed by former writers.

As may be seen from the section described above, *Pr. (Gigantella) primordialis* sp. n. Paeck. has been found in porous limestones resting on the horizon of green sandstones with *Sp. plenus* Hall. which is either the uppermost horizon of Tournai, or a transition horizon from Tournai to Visé. The overlying porous limestones should therefore be referred to the very base of Visé horizon, i. e. to the base of S_1 subzone. This fact is of special interest as being the only available literary evidence of such an extremely low statigraphic position of the genus *Gigantella* Sar.

---

* The manuscript of this work is in the hands of the Geological Survey Board.

The German *Gigantella* described by Paeckelman (9) may possibly be encountered also in the lower portion of Visé horizon, but it is a matter for regret that the author does not state its stratigraphic position with a precision that would be sufficient to connect it with the English subdivision of the Lower Carboniferous.

As regards the Moscow basin the absence of *Gigantella* in its lowest strata may be accounted for by the occurrence, in that basin, of the Continental Carboniferous series corresponding to S_1 subzone.

Moscow Institute of
Geological Survey.

Received
23. III. 1937.

REFERENCES

[1] С. М. Андронов, ИМЕН, сер. геолог., № 2 (1937). [2] М. А. Болховитинова, Новые данные по стратиграфии юго-восточной части 58 листа общей геологической карты, ОНТИ (1932). [3] В. Ф. Беспалов, Пробл. сов. геологии, 6, № 11 (1936). [4] А. Краснопольский, Геол. исслед. и развед. работы по линии Сиб. жел. дор., вып. 21 (1900). [5] К. И. Лисицын, Изв. Донск. политехн. ин-та, Новочеркасск, 13 (1929). [6] А. П. Ротай, Тр. ГГРУ, вып. 73 (1931). [7] Т. Г. Сарычева, Тр. геол. научно-исслед. института при физ.-мат. факультете 1 МГУ, вып. 1 (1928). [8] Y. Chao, Palaeontologia Sinica, 5, f. 2 (1927). [9] Paeckelman, Die Brachiopoden des Deutschen Untercarbons. 2. Teil: Die *Productionae* und *productus* ähnlichen *Chonetinae*, Berlin (1931). [10] A. Vaughan, Q. J. G. S., 61 (1905).

PRIX D'ABONNEMENT

Un an — 36 livraisons . . . . . $ 15
Six mois — 18 livraisons . . . $ 7.50
(Séparément $ 0.50 le numéro)
«Nouvelle Série» pour 1933 — 6 livraisons . . . $ 2
«Nouvelle Série» pour 1934 — l'année complète épuisée; volume III (9 livraisons) et volume IV (9 livraisons) en vente à $ 3 le volume.

URSS — Mezhdunarodnaja Kniga, 18, Kuznetskij Most, Moscou (Compte courant n° 263 Banque d'Etat de l'URSS, 12, Neglinnaja, Moscou).
NORTH and SOUTH AMERICA — BOOKNIGA CORPORATION. 255 Fifth Avenue. New York, N. Y.
GREAT BRITAIN — W. H. SMITH & son, Ltd. Strand House, Portugal St London, W. C. 2.
FRANCE — MESSAGERIES HACHETTE. Service Abonnements. 111, Rue Réaumur. Paris, 2e.
SUISSE — Administration du Journal „LE TRAVAIL". 39, Rue de Lausanne. Genève.
ROUMANIE — Deposit de Ziare Luvru. Calea Victoriei 30. Bucarest.
Č. S. R. — MELANTRICH, Akc. Spol. Knihkupectvi — odd. slovanskych knih. Vaclavské nam 42. Praha II. (Ucet post. spor. 20. 208).
BELGIQUE — LIBRAIRIE A.U.S. 38. Vieille Halle-aux-Blés. Bruxelles (C. C. P. 1861. 15 P. Driessen).
CHINA — AMERICAN BOOK & SUPPLY Co 841 Bubbling Well Road. P. O. Box 1798. Shanghai.
HOLLAND — J. GINSBERG. University Booksel er. Kort Rapenburg, 17. Leiden
LIETUVA — „MOKSLAS". Knygynas. Laisves Aleja, 46. Kaunas.
TURQUIE — LIBRAIRIE HACHETTE. Succursale de Turquie. 469, Avenue de l'Indépendance. Boite Postale Pera 2219. Istanbul.
POLSKA — GEBETHNER & WOLFF. Filja 1. Krakowskie Przedm. 15. Warszawa.
ROSSICA. Chmielna 5. Warszawa.
KSIEGARNIA GIRSZOWSKIEGO. Zamkowa 13. Wilno.
PALESTINE — SAFRAN Co. Jaffa Rd. 36. P. O. B. 223. Haifa.
BULGARIA — T. F. TSCHIPEFF. Buchhandlung. Boulevard Dondoukoff 20/22 Sofia. (Postscheckkonto-Sofia 540).
SVERIGE — C. E. FRITZE'S Kgl, Hofbuchhandlung. Fredsgatan, 2 Stockholm.
LATVIJA — A./S. VALTERS un RAPA. Teatra iela Nr. 11. Riga (Pasta tek. rek Nr. 675)
NORGE — JOHAN GRUNDT TANUM. Tidligere Aschehougs Boghandel. Karl Johans gt. 43. Oslo.

Les «Comptes Rendus», largement expédiés à l'étranger dans les centres scientifiques les plus importants et dans les bureaux de nombreuses revues, sont le seul organe scientifique en URSS, donnant tous les dix jours un compte rendu des plus récentes données de la science soviétique.

Prix — $ 1.00